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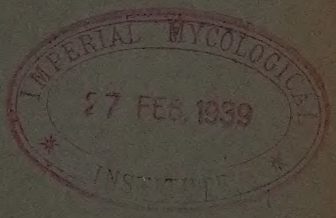

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MATHÉMATIQUES

SUR UNE FONCTIONNELLE DANS L'ESPACE DE HILBERT

Par A. G. PINSKER

(Présenté par I. Vinogradow, de l'Académie, le 21.VI.1938)

Dans cette note nous indiquons la forme générale d'une fonctionnelle continue dans L^2 (L étant l'espace des fonctions à carré sommables $x(t)$ $(a \leqslant t \leqslant b)$ et partiellement additive dans le sens que la relation d'additivité

$$\Phi(x+y) = \Phi(x) + \Phi(y)$$

est remplie pour tout couple d'éléments $x = x(t)$, $y = y(t)$ qui satisfait à la condition d'orthogonalité:

$$\int_a^b x(t)\, y(t)\, dt = 0.$$

Théorème 1. *La forme générale d'une fonctionnelle dans L^2 continue et additive pour les éléments orthogonaux est donnée par la formule*

$$\Phi(x) = \int_a^b [Ax^2(t) + x(t)\psi(t)]\, dt \quad (A = \text{const},\ \psi(t) \in L^2). \tag{1}$$

Démonstration. Il est facile de voir que la formule (1) définit réellement une fonctionnelle dans L^2 satisfaisant aux conditions du théorème. Soit, maintenant, $\Phi(x)$ une fonctionnelle de la forme considérée.

Considérons l'élément

$$x_e = x_e(t) = \begin{cases} \xi \text{ pour } t \in e \\ 0 \text{ pour } t \in ce \end{cases}$$

(e, étant un ensemble arbitraire de l'intervalle principal $[a, b]$) et posons

$$\Phi(x_e) = \varphi_e(\xi).$$

Soit

$$x^* = x_{e,e'}(t) = \begin{cases} \xi \text{ pour } t \in e \\ \alpha \text{ pour } t \in e' \\ 0 \text{ pour } t \in c(e+e') \end{cases} \quad \text{et} \quad y^* = y_{e,e'}(t) = \begin{cases} \eta \text{ pour } t \in e \\ -\alpha \text{ pour } t \in e' \\ 0 \text{ pour } t \in c(e+e') \end{cases}$$

où e et e' sont des ensembles mesurables sans points communs et où

$$\xi\eta \operatorname{mes} e - \alpha^2 \operatorname{mes} e' = 0.$$

Vu l'orthogonalité des éléments x^* et y^* et d'après l'additivité de la fonctionnelle

$$\Phi(x^* + y^*) = \Phi(x^*) + \Phi(y^*),$$

il résulte l'identité suivante qui lie les fonctions $\varphi_e(\xi)$ et $\varphi_e(\eta)$:

$$\varphi_e(\xi+\eta) = \varphi_e(\xi) + \varphi_e(\eta) + \varphi_{e'}(\alpha) + \varphi_{e'}(-\alpha). \tag{2}$$

1. Admettons que la fonctionnelle $\Phi(x)$ est impaire: $\Phi(-x) = -\Phi(x)$. Dans ce cas l'identité (2) prend la forme:

$$\varphi_e(\xi+\eta) = \varphi_e(\xi) + \varphi_e(\eta).$$

Il en suit que si $\bar{x}(t)$ et $\bar{y}(t)$ sont des fonctions à un nombre fini de valeurs distinctes $\bar{x}(t) = \xi_i$ et $\bar{y}(t) = \eta_i$ pour $t \in e_i$ $(i = 1, 2, \ldots n)$, l'on a

$$\Phi(\bar{x}+\bar{y}) = \sum_{i=1}^{n} \varphi_{e_i}(\xi_i+\eta_i) = \sum_{i=1}^{n} \varphi_{e_i}(\xi) + \sum_{i=1}^{n} \varphi_{e_i}(\eta_i) = \Phi(\bar{x}) + \Phi(\bar{y}).$$

Ainsi, la fonctionnelle $\Phi(x)$ est additive pour les fonctions à un nombre fini de valeurs distinctes et d'après sa continuité elle est additive en général et, par conséquent, est une fonctionnelle linéaire:

$$\Phi(x) = \int_a^b x(t)\psi(t)\,dt \qquad (\psi(t) \in L^2).$$

2. Admettons que la fonctionnelle $\Phi(x)$ est paire: $\Phi(-x) = \Phi(x)$. Posons dans l'identité (2) $\alpha = \eta - \xi$ et $\operatorname{mes} e = \operatorname{mes} e'$. On a alors

$$\varphi_{e'}(\xi) = \frac{1}{2}[\varphi_e(2\xi) - 2\varphi_e(\xi)]. \tag{3}$$

Analogiquement, si les ensembles e'' et e sont sans points communs et $\operatorname{mes} e'' = \operatorname{mes} e$, on a

$$\varphi_{e''}(\xi) = \frac{1}{2}[\varphi_e(2\xi) - 2\varphi_e(\xi)], \tag{3'}$$

c'est-à-dire

$$\varphi_{e'}(\xi) = \varphi_{e''}(\xi).$$

En particulier, pour deux intervalles d_1 et d_2 de la longueur $\frac{1}{n}$ nous aurons $\varphi_{d_1}(\xi) = \varphi_{d_2}(\xi)$. Si e_1 et e_2 sont deux systèmes d'intervalles aux bornes rationnelles, alors en les divisant en intervalles d'une longueur convenable $\frac{1}{n}$ nous trouverons, que

$$\frac{1}{\operatorname{mes} e_1}\varphi_{e_1}(\xi) = \frac{1}{\operatorname{mes} e_2}\varphi_{e_2}(\xi).$$

D'après la continuité on conclue qu'il en est de même pour tous les ensembles, et, par conséquent,

$$\frac{1}{\text{mes}\, e}\varphi_e(\xi) = A(\xi). \tag{4}$$

De la relation (3) nous avons

$$\varphi_e(2\xi) = 2\varphi_e(\xi). \tag{5}$$

Posons, maintenant, dans l'identité (2) $\xi = \eta = 1$, alors

$$\text{mes}\, e = \alpha^2 \,\text{mes}\, e' \quad \text{et} \quad \varphi_e(2) = 2\varphi_e(1) + 2\varphi_{e'}(\alpha),$$

ou en tenant compte de (4) et (5), on aura

$$\varphi_{e'}(\alpha) = \varphi_e(1) = A(1)\,\text{mes}\, e = A\,\alpha^2\,\text{mes}\, e', \qquad (A(1) = A).$$

Soit $\bar{x}(t)$ une fonction à un nombre fini de valeurs distinctes: $\bar{x}(t) = \alpha$ pour $t \in e_i$ $(i = 1, 2, \ldots n)$.

On a

$$\Phi(\bar{x}) = \sum_{i=1}^{n} \varphi_{e_i}(\alpha_i) = A \sum_{i=1}^{n} \alpha_i^2 \,\text{mes}\, e_i = A \int_a^b \bar{x}^2(t)\, dt$$

et d'après la continuité de la fonctionnelle

$$\Phi(x) = A \int_a^b x^2(t)\, dt,$$

où $x = x(t)$ est un élément arbitraire de l'espace L^2.

3. De même que la fonctionnelle $\Phi(x)$, les fonctionnelles $\Phi_1(x) = \frac{1}{2}[\Phi(x) + \Phi(-x)]$ et $\Phi_2(x) = \frac{1}{2}[\Phi(x) - \Phi(-x)]$ seront de la même forme; mais comme la fonctionnelle $\Phi_1(x)$ est paire, $\Phi_2(x)$ est impaire et $\Phi(x) = \Phi_1(x) + \Phi_2(x)$, on aura

$$\Phi(x) = \int_a^b [Ax^2(t) + x(t)\,\psi(t)]\, dt,$$

c. q. f. d.

Les raisonnements précédents sont facilement généralisés dans le cas, où les valeurs de l'opération appartiennent à quelque espace abstrait et linéaire Y. Alors le théorème suivant a lieu.

Théorème 1. *La forme générale d'une opération continue et additive pour les éléments orthogonaux transformant L^2 dans Y est donnée par la formule*

$$\Phi(x) = y_0 \int_a^b x^2(t)\, dt + U(x),$$

où $y_0 \in Y$ et $U(x)$ étant une opération linéaire.

Vu l'isométrie des espaces L^2 et l^2, on peut énoncer les théorèmes analogues pour l^2:

Théorème 2. *La forme générale d'une fonctionnelle continue dans* l^2, *additive pour les couples d'éléments orthogonaux, est donnée par la formule*

$$\Phi(x)=\sum_{i=1}^{\infty}(A\,\xi_i^2+a_i\,\xi_i),$$

où

$$x=(\xi_1,\xi_2,\ldots),\quad A=\text{const},\quad \sum_{i=1}^{\infty}a_i^2<+\infty.$$

Théorème 2'. *La forme générale d'une opération continue et additive pour les éléments orthogonaux transformant* l^2 *dans* Y *est donnée par la formule*

$$\Phi(x)=y_0\sum_{i=1}^{\infty}\xi_i^2+U(x),$$

où $y_0\in Y$ *et* $U(x)$ *étant une opération linéaire.*

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ASTROPHYSICS

ON THE GRAVITATIONAL INSTABILITY OF A MATERIAL SPHERE

By A. B. SEVERNY

(Communicated by V. G. Fessenkov, Member of the Academy, 24. VI. 1938)

The conception of gravitational instability has been put forward by Jeans (1) in connection with different problems of cosmogonical character. Consider a sphere of radius a containing «compressible» matter obeying the laws of hydrodynamics, so that the following equations of motion hold:

$$f_x = X - \frac{1}{\rho}\frac{\partial p}{\partial x}, \quad \text{etc.}, \tag{1}$$

where f_x, etc., are the acceleration components, X, etc., are the corresponding components of the external force, p is the pressure and ρ is the density of matter at the point x, y, z. Let us consider a varied motion such that the centre of the element $dx\,dy\,dz$ is at the time t at $x+\xi,\ y+\eta,\ z+\zeta$ (or, let us consider another configuration for which the values p, ρ, etc., refer not to the point x, y, z, but to the point $x+\xi,\ y+\eta,\ z+\zeta$). Let us assume that

1) there exists the potential U of the external force,

2) the pressure p depends on the density ρ only.

By comparison of the varied configuration with the original one we obtain the equation determining the virtual displacement

$$\bar{s} = \bar{i}\xi + \bar{j}\eta + \bar{k}\zeta$$

(loc. cit.):

$$\frac{\partial^2 \sigma}{\partial t^2} - \operatorname{div}\bar{A} = 4\pi G\rho\sigma + \nabla^2\left(\frac{\partial p}{\partial \rho}\sigma\right), \tag{2}$$

where

$$\sigma = \operatorname{div}\bar{s},\ \operatorname{div}\bar{A} = \frac{\partial}{\partial x}(\bar{s},\ \operatorname{grad} f_x) +$$

$$+ \frac{\partial}{\partial y}(\bar{s},\ \operatorname{grad} f_y) + \frac{\partial}{\partial z}(\bar{s},\ \operatorname{grad} f_z),$$

i. e. σ is a condensation (or rarefaction), connected with the virtual displacement $\bar{s}$. We shall call the state of matter gravitationally instable in the sense of Jeans (or J-instable), if the displacements ξ, η, ζ (and consequently σ) are monotonous functions of t. If these quantities are periodic functions of t, the configuration will be J-stable (but it may be instable according to Jacobi's and Liapounoff's criteria).

Further, let us assume that 3) the acceleration $\bar{f}$ does not depend on x, y, z and t (i. e. there are no systematical motions of the matter; in particular, the matter is at rest). Then $\operatorname{div} \bar{A} = 0$, and the fundamental equation (2) becomes

$$\frac{\partial^2 \sigma}{\partial t^2} = 4\pi G \rho \sigma + \nabla^2 \left(\frac{\partial p}{\partial \rho} \sigma \right). \tag{3}$$

In accordance with the second assumption we suppose that the state of matter is determined by the most general law

$$p = K \rho^\gamma \qquad (\gamma > 0), \tag{4}$$

and assume, further, that the density satisfies the following condition:

$$\rho_{\max} > \rho > \rho_{\min} \qquad (> 0). \tag{5}$$

Changing to polar coordinates, we take the boundary conditions for σ in the form

$$\left.\begin{aligned} &\text{(a)} \quad r = 0, \quad \sigma \text{ and } \frac{\partial \sigma}{\partial r} \text{ are finite and continuous,} \\ &\text{(b)} \quad r = a, \quad \sigma(a) = 0. \end{aligned}\right\}$$

The solution of (3) we assume to be of the form

$$\sigma = e^{ikt} f(r)\, Y(\theta, \varphi).$$

Under the conditions of continuity of Y on the sphere $r = 1$, Y will be a spherical harmonic of order l, and for $f(r)$ we obtain

$$\frac{1}{r^2} \frac{d}{dr} \left[r^2 \frac{d}{dr} \left(\frac{\partial p}{\partial \rho} f(r) \right) \right] + \left[m(r) - \frac{l(l+1)}{r^2} \right] \frac{\partial p}{\partial \rho} f(r) = 0$$

or putting

$$u(r) = \frac{\partial p}{\partial \rho} r f(r),$$

we obtain for u the fundamental Sturm-Liouville equation for the problem

$$\frac{d^2 u}{dr^2} - q(r)\, u + \lambda s(r)\, u = 0, \tag{7}$$

where

$$s(r) = \frac{1}{K\gamma} \rho^{1-\gamma}, \quad q(r) = \frac{l(l+1)}{r^2} - \frac{4\pi G}{K\gamma} \rho^{2-\gamma}, \quad \lambda \equiv k^2. \tag{8}$$

λ is an «Eigenwert» of the problem. According to the above definition of J-instability, the mass of matter will be J-instable, if there exists one λ, $\lambda_{\min}$, say, with

$$\lambda_{\min} < 0. \tag{9}$$

Our aim is to establish necessary and sufficient conditions for J-instability, or necessary and sufficient conditions for the existence of negative «Eigenwerts» of the equation (7) with the boundary conditions (6) (a) and (b), which for the function u become, by (5),

$$u(0) = u(a) = 0. \tag{10}$$

In virtue of (10) we have

$$\lambda_n = \int_0^a [u_n'^2 + q(r)\, u_n^2]\, dr, \tag{11}$$

where u_n is the corresponding «Eigenfunktion». The necessary condition for the inequality (9) to be satisfied is: in the interval $(0, a)$ there should exist a subinterval (ξ, a) in which the function $q(r)$ is negative, i. e.

$$q(r) = \frac{l(l+1)}{r^2} - \frac{4\pi G}{K\gamma}\rho^{2-\gamma} < 0, \qquad 0 < \xi \leqslant r \leqslant a,$$

or

$$\frac{l(l+1)}{a^2} > \frac{4\pi G}{K\gamma}(\rho^{2-\gamma})_{\min},$$

since $\frac{l(l+1)}{r^2}$ is a monotonous function. Therefore a necessary condition for one of the numbers λ to be negative is

$$a > \sqrt{l(l+1)\frac{K\gamma}{4\pi G\,(\rho^{2-\gamma})_{\min}}} \equiv a_c, \tag{12}$$

A similar condition may be also found from principal propositions of Sturm's theory. From (11) we see that for sufficiently large l no λ will be negative, consequently only the lowest vibrations are J-instable.

For the establishment of a sufficient condition we shall apply a theorem of Courant([2]), according to which the «Eigenwert» λ_n increases, if in the whole interval $(0, a)$ $q(r)$ is increased and $s(r)$ is diminished. We have

$$q(r) < q_M(r) = \frac{l(l+1)}{r^2} - \frac{4\pi G}{K\gamma}(\rho^{2-\gamma})_{\min},$$

$$s(r) > s_m(r) = \frac{1}{K\gamma}(\rho^{1-\gamma})_{\min}$$

and consequently every «Eigenwert» of the equation

$$\frac{d^2u}{dr^2} - q_M u + \lambda^* s_m u = 0 \tag{13}$$

is greater than the corresponding «Eigenwert» of our problem.

Putting

$$m = \lambda^* s_m + \frac{4\pi G}{K\gamma}(\rho^{1-\gamma})_{\min}, \tag{14}$$

we shall have instead of (13)

$$\frac{d^2u}{dr^2} + \left(m - \frac{l(l+1)}{r^2}\right)u = 0,$$

the «Eigenfunktions» of which satisfying the condition $u(0) = 0$ are

$$u_l(r) = \sqrt{r}\, J_{l+\frac{1}{2}}(\sqrt{m}r),$$

where $J_{l+\frac{1}{2}}$ is Bessel's function of order $l+\frac{1}{2}$. The «Eigenwert's» λ^* are determined by the condition $u(a) = 0$, whence

$$\sqrt{m_{il}}\,a = \pm q_{il}, \qquad i = 0, 1, 2, \ldots,$$

where q_{il} are the successive roots of Bessel's function of order $l+\frac{1}{2}$. By (14) we have

$$\lambda^*_{il} = \left[\left(\frac{q_{il}}{a}\right)^2 - \frac{4\pi G}{K\gamma}(\rho^{2-\gamma})_{\min}\right]\frac{1}{s_{\min}},$$

and λ_{il}^{*} will be negative, if there exists such a q_{il} that

$$a > q_{il} \sqrt{\frac{K\gamma}{4\pi G\,(\rho^{2-\gamma})_{\min}}}.$$

The lowest value q_{il} is $q_{00}\,(=3.172)$; therefore, if the condition

$$a > 3.172 \sqrt{\frac{K\gamma}{4\pi G\,(\rho^{2-\gamma})_{\min}}} \equiv a_c \qquad (15)$$

is satisfied, at least the lowest «Eigenwert» will be negative; but since $\lambda < \lambda^{*}$, λ_{00} will be also negative. Therefore, the condition (15) is a sufficient condition for the J-instability. The comparison of the necessary condition (12) with the sufficient condition (15) (for $l=1$) shows that a_c exceeds a_0 by 3.2 times.

Our general conclusion is as follows: the gravitational instability (in the above sense) of a material sphere at the state of matter described by (4) and with the density satisfying to (5) appears when the radius of this sphere exceeds a certain critical value a_c. A detailed analysis of special cases of J-instability will be the subject of further communications. However, we may point out already here that taking $1 < \gamma < 2$ and $\rho_{\min} \approx 10^{-30}$ g/cm^3 (the density of matter in space) we obtain for a_c values of the order of 1—100 parsecs, i. e. values somewhat lower than the dimensions of the gallactic systems.

Physico-Mathematical Association of the
Academy of Sciences of the USSR.

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REFERENCES

[1] Jeans, Astron. a. Cosmogony, 337—341 (1928). [2] R. Courant und D. Hilbert, Methoden d. mathem. Physik., I, Kap. VI, § 2 (1931).

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MÉCANIQUE

SUR LA CONSTRUCTION DES RÉGIONS DE LA STABILITÉ AU SENS DE LIAPOUNOFF

Par N. MOÏSSEIEV

(Présenté par V. G. Fessenkoff, de l'Académie, le 24. VI. 1938)

Soit

$$\dot{x}_s = X_s(x_1, x_2, x_3, \ldots x_n) \quad (s = 1, 2, \ldots n) \tag{1}$$

un système des équations différentielles et soit

$$P^0(x_1^0, x_2^0, \ldots x_n^0) \tag{2}$$

un point de l'espace des phases, se mouvant de manière que ses coordonnées vérifient les équations

$$\dot{x}_s^0 = X_s(x_1^0, x_2^0, \ldots x_n^0) = X_s^0 \quad (s = 1, 2, \ldots n). \tag{3}$$

En utilisant les désignations

$$\xi_s = x_s - x_s^0, \tag{4}$$

nous mettons les équations (1) du mouvement troublé (Liapounoff) sous la forme:

$$\dot{\xi}_s = \Xi\ (\xi_1, \xi_2, \ldots \xi_n, x_1^0, x_2^0, \ldots x_n^0) \quad (s = 1, 2 \ldots n). \tag{5}$$

Les seconds membres de ces équations sont des certaines fonctions bien définies de ξ_s qui deviennent toutes nulles lorsque tous les ξ_s s'annulent simultanément. Nous supposons que Ξ_s sont des fonctions bornées, uniformes et continues des variables ξ_s et x_s^0 partout dans l'espace des phases $(x_1^0, x_2^0, \ldots x_n^0)$, sauf, peut être, pour les points singuliers dont nous allons faire exception, de même que pour ses voisinages immédiats, dans tous ce qui va suivre.

Soit

$$V = V(\xi_1, \xi_2, \ldots \xi_n, x_1^0, x_2^0, \ldots x_n^0) \tag{6}$$

une fonction de Liapounoff [1] par rapport aux variables ξ_s; elle est, par conséquent, une fonction uniforme et continue des ξ_s, s'annulant lorsque la valeur de

$$\rho = \left| \sqrt{\sum_{s=1}^{n} \xi_s^2} \right| \tag{7}$$

devient nulle, quelles que soient les valeurs des variables x_s^0. Nous supposons aussi qu'à tout nombre positif ε on peut faire correspondre un nombre positif δ, indépendant des variables x_s^0, tel que nous aurons dans tout l'espace des phases $(x_1^0, x_2^0, \ldots x_n^0)$:

$$|V| < \varepsilon \tag{8}$$

chaque fois que la seule condition

$$\rho < \delta \tag{9}$$

sera remplie.

Cette dernière supposition nous garantit que sur chaque semitrajectoire positive T_+^0, correspondant à un ensemble arbitraire des conditions initiales et à une solution particulière (déterminée par ces conditions) (2) des équations (1), la fonction (6), dans laquelle les variables x_s^0 seront remplacées par ses expressions explicites en fonction du temps t, sera une fonction de Liapounoff, qui admet une limite supérieure infiniment petite (2).

Envisageons une certaine position fixe du point P^0 (2) et formons dans l'espace $(\xi_1, \xi_2, \ldots \xi_n)$ trois ensembles des points suivants:

1) V^0, en chaque point duquel nous aurons $V = 0$ *,

2) V^+, en chaque point duquel nous aurons $V > 0$, et enfin

3) V^-, en chaque point duquel nous aurons $V < 0$.

Désignons respectivement par $\rho(V^0)$, $\rho(V^+)$ et $\rho(V^-)$ les distances du point P^0 aux ensembles V^0, V^+ et V^-. Ces trois distances seront les fonctions uniformes et non négatives du point P^0 dans l'espace des phases $(x_1^0, x_2^0, \ldots x_n^0)$. En utilisant ces fonctions nous pouvons construire dans l'espace des phases $(x_1^0, x_2^0, \ldots x_n^0)$ quatre ensembles des points suivants:

1) $\rho^+(V^0, V^+)$, en chaque point duquel $\rho(V^0)$, aussi bien que $\rho(V^+)$, est positif,

2) $\rho^+(V^0, V^-)$, en chaque point duquel $\rho(V^0)$ et $\rho(V^-)$ sont positifs.

3) $\rho^0(V^+)$, en chaque point duquel $\rho(V^+)$ est nul, et enfin

4) $\rho^0(V^-)$, en chaque point duquel $\rho(V^-)$ est nul.

Formons maintenant à l'aide des équations (5) et (3) la dérivée totale $\dot{V}$ de la fonction V par t. Nous obtenons:

$$\dot{V} = \sum_{s=1}^{n} \frac{\partial V}{\partial \xi_s} \Xi_s + \sum_{s=1}^{n} \frac{\partial V}{\partial x_s^0} X_s^0. \tag{10}$$

En vertu des propriétés des fonctions de Liapounoff ** ainsi qu'en vertu des suppositions faites par nous relativement aux équations (3) et (5), la dérivée (10) sera une fonction uniforme et continue des variables ξ_s et x_s^0. Nous construisons dans l'espace $(\xi_1, \xi_2, \ldots \xi_n)$, correspondant à une position fixe quelconque du point (2), trois ensembles des points suivants:

1) $\dot{V}^0$, en chaque point duquel on a $\dot{V}^0 = 0$ ***,

---

* Quant au point P^0, dans lequel nous avons toujours $V = 0$, nous ne le comptons pas parmi les points de l'ensemble V^0.

** Il n'est pas nécessaire de supposer que les dérivées partielles soient continues (3). La théorie qui sera exposée peut subsister aussi bien sans cette supposition.

*** Le point P^0 n'est pas compté parmi les points de l'ensemble $\dot{V}^0$.

2) V^+, en chaque point duquel on a $V > 0$, et enfin

3) $\dot{V}-$, en chaque point duquel on a $\dot{V} < 0$.

D'après cela nous pouvons trouver trois fonctions uniformes et non négatives du point (2): $\rho(\dot{V}^0)$, $\rho(\dot{V}^+)$ et $\rho(\dot{V}^-)$ représentant respectivement les distances du point (2) aux ensembles $\dot{V}^0$, $\dot{V}^+$ et $\dot{V}^-$.

En utilisant ces fonctions nous pouvons construire dans l'espace des phases $(x_1^0, x_2^0, \ldots x_n^0)$ quatre ensembles des points suivants:

1) $\rho^+(\dot{V}^0, \dot{V}^+)$, en chaque point duquel $\rho(\dot{V}^0)$ et $\rho(\dot{V}^+)$ sont positifs,

2) $\rho^+(\dot{V}^0, \dot{V}^-)$, en chaque point duquel $\rho(\dot{V}^0)$ et $\rho(\dot{V}^-)$ sont positifs,

3) $\rho^+(\dot{V}^+)$, en chaque point duquel $\rho(\dot{V}^+)$ est positif, et enfin

4) $\rho^+(\dot{V}^-)$, en chaque point duquel $\rho(\dot{V}^-)$ est positif.

D'après cela formons l'intersection des ensembles:

$$(S)_{V^-} = \rho^+(V^0, V^+)\rho^+(\dot{V}^-), \tag{11}$$

ainsi que l'intersection

$$(S)_{V^+} = \rho^+(V^0, V^-)\rho^+(\dot{V}^+). \tag{12}$$

Appelons chacun de ces deux ensembles des points dans l'espace des phases $(x_1^0, x_2^0, \ldots x_n^0)$ — région de la stabilité au sens de Liapounoff, construite à l'aide de la fonction (6), et désignons chacun de ces deux ensembles (11) et (12) par $(S)_V$*.

Formons aussi les intersections

$$(IS)_{V^-} = \rho^0(V^-)\rho^+(\dot{V}^0, V^+), \tag{13}$$

et

$$(IS)_{V^+} = \rho^0(V^+)\rho^+(\dot{V}^0, \dot{V}^-). \tag{14}$$

Chacun de ces deux ensembles (13) et (14) sera appelé par nous — région de l'instabilité au sens de Liapounoff, établie à l'aide de la fonction (6), et sera désigné par $(IS)_V$,

Quant aux intersections

$$(As)_{V^-} = \rho^+(V^0, V^+)\rho^+(\dot{V}^0, \dot{V}^-) \tag{15}$$

et

$$(As)_{V^+} = \rho^+(V^0, V^-)\rho^+(\dot{V}^0, \dot{V}^+), \tag{16}$$

nous appelons chacune d'elles — région de la stabilité asymptotique au sens de Liapounoff, établie à l'aide de la fonction (6), et nous la désignerons par $(As)_V$**.

Formons enfin les intersections

$$(ISa)_{V^-} = \rho^+(V^0, V^+)\rho^+(\dot{V}^0, \dot{V}^+) \tag{17}$$

et

$$(ISa)_{V^+} = \rho^+(V^0, V^-)\rho^+(\dot{V}^0, \dot{V}^-). \tag{18}$$

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* Si nous voulions nous borner à la construction de cette région de la stabilité $(S)_V$ seulement, nous pourrions dans ce cas omettre la condition liée avec les inégalités (8) et (9).

** Cette région n'existe pas, si nous avons $\sum_{s=1}^{n} \frac{\partial X_s}{\partial x_s} \equiv 0$ pour toutes les valeurs des x_s.

Chacun de ces deux ensembles des points sera appelé par nous — région de l'instabilité absolue au sens de Liapounoff, établie à l'aide de la fonction (6), et sera désigné par $(ISa)_V$ **.

Pour être bref dans ce qui va suivre, nous dirons que la semitrajectoire positive T_+^0 est enfoncée dans une région $f > 0$ d'une manière uniforme par rapport au f. Cela doit signifier que tous les points de la semitrajectoire T_+^0 sont situés à l'intérieur de la région $f > 0$ et que la limite inférieure précise des valeurs de f sur T_+^0 est distincte de zéro.

Alors nous pouvons énoncer quatre propositions suivantes, révélant le sens des régions dont la construction est exposée plus haut.

1. Chaque solution particulière du système (1), à laquelle correspond la semitrajectoire positive T_+^0, enfoncée dans la région $(S)_{V^-}$ [ou bien $(S)_{V^+}$] d'une manière uniforme par rapport aux $\rho(V^0)$, $\rho(V^+)$, $(\dot{V}^-)$ [ou bien $\rho(V^0)$, $\rho(V^-)$, $\rho(\dot{V}^+)$] doit être stable au sens de Liapounoff.

Cette affirmation est une conséquence immédiate du théorème 1 de la deuxième méthode de Liapounoff (4).

2. Chaque solution particulière du système (1), à laquelle correspond la semitrajectoire T_+^0, enfoncée dans la région $(IS)_{V^-}$ [ou bien $(IS)_{V^+}$] d'une manière uniforme par rapport aux $\rho(\dot{V}^0)$, $\rho(\dot{V}^+)$ [ou bien $\rho(\dot{V}^0)$, $\rho(\dot{V}^-)$], doit être instable au sens de Liapounoff.

Cette affirmation n'est qu'une simple conséquence du deuxième théorème de Liapounoff (5).

3. Chaque solution particulière, à laquelle correspond la semitrajectoire T_+^0, enfoncée dans la région $(As)_{V^-}$ [ou bien $(As)_{V^+}$] d'une manière uniforme par rapport aux $\rho(V^0)$, $\rho(V^+)$, $\rho(\dot{V}^0)$, $\rho(\dot{V}^-)$ [ou bien $\rho(V^0)$, $\rho(V^-)$, $\rho(\dot{V}^0)$, $\rho(\dot{V}^+)$], doit posséder la stabilité asymptotique au sens de Liapounoff.

Cette affirmation est une conséquence de la deuxième remarque du premier théorème de Liapounoff (6).

4. Chaque solution particulière du système (1), à laquelle correspond la semitrajectoire T_+^0, enfoncée dans la région $(ISa)_{V^-}$ [ou bien $(ISa)_{V^+}$] d'une manière uniforme par rapport aux $\rho(V^0)$, $\rho(V^+)$, $\rho(\dot{V}^0)$, $\rho(\dot{V}^+)$ [ou bien $\rho(V^0)$, $\rho(V^-)$, $\rho(\dot{V}^0)$, $\rho(\dot{V}^-)$], doit posséder l'instabilité absolue au sens de Liapounoff.

L'expression «l'instabilité absolue» étant entendue au sens de M. Doubochin, cette affirmation n'est qu'une conséquence évidente du deuxième théorème de Liapounoff (5).

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LITTÉRATURE CITÉE

[1] A. Liapounoff, Ann. de la Faculté des Sciences de Toulouse, 2-me sér., 9, 256 (1907). [2] A. Liapounoff, ibid., p. 258. [3] N. Moïsseiev, C. R. Acad. Sci. URSS, I (X), № 4 (1936). [4] A. Liapounoff, loc. cit., p. 259. [5] A. Liapounoff, ibid., p. 262. [6] A. Liapounoff, ibid., p. 261.

* Voir la remarque précédente.

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MÉCANIQUE

SUR LES RÉGIONS DE LA STABILITÉ DANS L'ESPACE DES PHASES

Par N. MOÏSSEIEV

(Présenté par V. G. Fessenkoff, de l'Académie, le 24.VI.1938)

Dans la note précédente (1) nous avons établi un algorithme bien défini, à l'aide duquel on peut construire les régions dans l'espace des phases, que nous avons appelées—régions de la stabilité, régions de l'instabilité et ainsi de suite. Ici nous voulons exposer quelques remarques complémentaires concernant la question de l'importance des régions mentionnées pour l'analyse générale qualitative des problèmes dynamiques, ainsi que de la possibilité de certaines généralisations de la méthode de la construction de ces régions. Les désignations de la note citée sont à conserver.

La propriété la plus essentielle des régions en question est la suivante: s'ils existent parmis les semitrajectoires T_+^0 du problème (1) telles, qu'elles soient uniformement enfoncées dans une des régions en question, le mouvement le long de ces semitrajectoires est necessairement stable au sens de Liapounoff, si la région appartient au type S, instable au même sens, si la région appartient au type IS, et ainsi de suite.

À cause de cela le problème de la construction des régions de la stabilité etc. doit être reconnu intimement lié au problème qui est d'une importance extrême pour toute la théorie qualitative des équations différentielles, à savoir au problème de la séparation de l'ensemble total des mouvements possibles dans un certain problème dynamique en sous-ensembles, chacun d'eux n'embrassant que des mouvements stables*, ou des mouvements instables, ou des mouvements asymptotiquement stables et ainsi de suite. Dans certains cas la pratique exige de ne distinguer que des trajectoires asymptotiquement stables, n'attachant pas l'attention à tout l'ensemble des trajectoires possibles. Ceci fait attribuer au dernier problème une importance assez grande au point de vue des appliations pratiques.

Le procédé de la construction des régions de la stabilité, exposé dans la note précédente, est un procédé purement qualitatif et n'exige aucunement l'intégration quantitative des équations différentielles du mouvement. Les régions, obtenues par ce procédé, sont en liaison intime avec les sous-ensembles des mouvements monotypiques au point de vue de la théorie de la stabilité, et on peut indiquer une série d'exemples, quand notre méthode donne avec simplicité des régions dans l'espace des phases, formées par les

* Ce problème a été posé pour la première fois par N. Četaïev (2,3) dans ses recherches approfondies sur les trajectoires stables en dynamique. Le sujet principal de l'analyse de Četaïev consiste non dans la recherche des régions, formées par les points de toutes les trajectoires stables, mais dans la recherche des trajectoires stables individuelles.

points des semitrajectoires monotypiques au point de vue de la théorie de Liapounoff.

Néanmoins, dans un cas arbitraire, une région obtenue par notre méthode peut ne pas contenir à son intérieur de semitrajectoires qui y soient uniformément enfoncées. La méthode en question ne nous fournit aucune possibilité de distinguer les cas, où les semitrajectoires uniformément enfoncées dans la région de stabilité ou de l'instabilité existent ou non. Cela exige une autre méthode quelconque, auxiliaire de l'analyse qualitative. Dans ce but peut être choisie la méthode de la localisation des trajectoires des phases (4).

D'autre part, on peut indiquer d'autres qnestions présentant également de l'intérêt, où la méthode de la construction des régions de la stabilité peut être employée comme méthode auxiliaire, combinée avec la méthode de la localisation. Il s'agit dans ces questions de localiser des trajectoires distinctes de phases et leurs segments finis. Alors la méthode de construction des régions de la stabilité peut servir à reconnaître les parties de l'espace des phases, où l'on peut compter sur l'heureuse application de la méthode de localisation.

Quant à l'application de la méthode de la construction des régions de stabilité, remarquons d'abord que les régions, que fournit notre méthode, dépendent beaucoup du choix de la fonction (6). En faisant varier d'une manière convenable la forme de cette fonction, nous pouvons obtenir à l'aide de notre méthode des régions de stabilité ou d'instabilité de construction heureuse et de l'étendue suffisante. Dans les cas les plus simples on peut se servir au lieu de la fonction (6) des formes quadratiques par rapport aux variables ξ_s, dont les coefficients dépendent de variables x_s^0, ou bien sont constants. Ce cas particulier de notre méthode est exposé avec certains détails dans un autre article (5).

Aux questions techniques on peut rapporter aussi la remarque suivante.

Dans tout ce que nous avons exposé nous n'avons employé que trois théorèmes de Liapounoff, représentant trois variantes des conditions, suffisantes pour la stabilité ou bien pour l'instabilité. En faisant usage des autres conditions suffisantes*, nous pourrions sans doute simplifier et perfectionner notre méthode.

Prêtons maintenant notre attention au problème de la généralisation possible de la théorie des régions de la stabilité ou de l'instabilité.

Premièrement, est à souligner la possibilité de la généralisation de ladite théorie aux cas où les seconds membres des équations (1) ainsi que la fonction (6) dépendent du temps t d'une manière explicite.

Deuxièmement, il faut noter qu'on peut également poser des questions analogues relativement aux régions de la stabilité o r b i t a l e, aux régions de la stabilité c o n d i t i o n n e l l e ou bien de la stabilité par rapport aux certaines fonctions des coordonnées des phases. Quant aux régions de la stabilité orbitale, cette question a été en partie examinée dans l'article déjà cité (5).

Troisièmement, remarquons qu'il est possible de construire les régions de la stabilité non seulement au sens de Liapounoff, mais aussi au sens des définitions de la stabilité du deuxieme, du troisième et du quatrième type, proposées par nous dans un de nos articles précédents (7), ou bien au sens de la définition de la stabilité du troisième type généralisé (8). La plus importante de ces questions se rapporte sans doute à la stabilité du troisième type (ou bien à la stabilité au sens de N. Kryloff et N. Bogoliouboff). Il est aussi possible de poser la question sur la construction des régions de l'anti-

* Par exemple, l'élégant théorème sur l'instabilité de N. Četaïev (8).

stabilité des divers types connus ([7],[8]). Enfin une théorie analogue peut être construite aussi pour les régions de la valeur constante de la probabilité de la stabilité et de l'antistabilité de divers types ([8],[9]).

À la fin de notre article nous nous permettons d'ajouter une remarque concernant la littérature de la question.

À notre connaissance, nos notes posent le problème de la construction des régions de la stabilité au sens de Liapounoff pour la première fois. Mais il faut indiquer aussi que l'auteur de ces notes était au courant d'un article de V. Stepanoff, publié en 1936 et consacré à la question de la construction des régions de la stabilité au sens de Jacobi des trajectoires d'un problème dynamique aux deux degrés de liberté ([10]). Aussi différents qu'ils soient, notre note présente et l'article de V. Stepanoff ont néanmoins pour base la même idée principale. Cette idée consiste notamment dans la représentation des conditions de la stabilité non comme des conditions dépendant du temps*, mais comme des conditions dépendant des coordonnées de phases du point dans son mouvement non troublé. Cette idée a permis à V. Stepanoff d'envisager la caractéristique de la stabilité Jacobienne comme une fonction du point dans l'espace des phases de Birkhoff et, par conséquent, comme une caractéristique régionale dans le même espace.

Nous avons considéré aussi dans nos articles la caractéristique de la stabilité au sens de Liapounoff comme une fonction du point de l'espace des phases et, par conséquent, comme une caractéristique régionale qualitative dans le même espace.

C'est le seul point commun entre l'article de V. Stepanoff et le nôtre.

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LITTÉRATURE CITÉE

[1] N. Moïsseiev, C. R. Acad. Sci. URSS, XX, № 6 (1938). [2] Н. Г. Четаев, Сборник Казанского авиац. ин-та, № 5, 3 (1936). [3] Н. Г. Четаев, Уч. зап. Каз. ун-та, **91**, № 4,3 (1931). [4] N. Moïsseiev, C. R. Acad. Sci. URSS, XVII, № 6 (1937). [5] Н. Моисеев, Тр. ГАИШ, **9**, вып. 2 (1938). [6] N. Četaïev, C. R. Acad. Sci. URSS, № 9 (1934). [7] Н. Моисеев, Тр. Воен. возд. академии, вып. 26 (1938). [8] N. Moïsseiev, C. R. Acad. Sci. URSS, XVI, № 6 (1937). [9] N. Moïsseiev, Math. ZS., **42**, 515 (1937). [10] В. Степанов, Астр. журн., **13**, 435 (1936).

* Ou bien d'une autre variable quelconque, dont on peut se servir en déterminant la position du point dans sa course non perturbée.

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PHYSICS

MEASUREMENT OF e/m_0 FOR β-PARTICLES OF RaC

By A. ALICHANOV, A. ALICHANIAN and M. KOZODAEV

(Communicated by A. F. Joffe, Member of the Academy, 25.VI.1938)

I. E. Tamm has advanced the hypothesis that, if electrons of a mass several times (2—5) exceeding that of an ordinary electron are present in cosmic radiation, then it is to be expected that electrons of the same kind will be emitted during β-disintegration too.

The ratio of the number of heavy electrons of a mass k times greater than the electron mass to the number of ordinary electrons will be defined by the following equation:

$$n = \left(\frac{W_0 - \sqrt{k^2 + p^2}}{W_0 - \sqrt{1 + p^2}} \right)^{2a}, \qquad (1)$$

where, according to Fermi, $a=1$ and, according to Uhlenbeck and Konopinsky, $a=2$. It is evident that n will be the greater, the greater the upper limit of the β-spectrum W_0 and the smaller the momentum of β-particle p.

Starting from the above considerations we measured e/m_0 for β-particles of RaC within the Hg∼1200—3000 region. The object of the present investigation was not an accurate measurement of e/m_0*, but we intended to establish the presence of heavy electrons even in the case if their number were only some tenths per cent of the number of ordinary electrons.

In a usual magnetic spectrograph in which electrons were focussed by the Danisz method, two plates of a condenser were placed, which were bent to a radius of 50 cm. The direction of the electric field was parallel to that of the magnetic field. The force acting on the electron throughout its motion in the condenser remained perpendicular to its trajectory, resulting in the electron motion in a circle of 50 cm radius. Simultaneously, the magnetic field was deflecting the electron within a circle of 10 cm radius, but in a plane perpendicular to the direction of the electric field. In the first part of its path from the source to the counter the electron was subjected to a simultaneous action both of the magnetic and the electric fields parallel to each other, whereas during the second part of the path it was subjected to the action of the magnetic field alone. The above method of measuring the e/m_0 has the advantage of maintaining the magnetic focussing, i. e. it makes it possible to have a great aperture ratio.

* An accurate measurement of e/m_0 for β-particles of RaE was carried out by Zahn and Spis, Phys. Rev., **53**, 365 (1938).

The scattering of electrons from the plates of the condenser hinders the solution of the problem. The plates were therefore placed 6 mm apart and slits were made in the condenser to cut off a beam 2 mm wide only. Owing to this fact, the plates of the condenser were little, if at all, irradiated by the straight beam of electrons. Moreover, the electrons scattered from the plates could be admitted into the counter only after multiple scattering.

According to equation (1), if W_0=3200 ekV and Hg=1500

		if a=1	if a=2
for	k=2	n=0.97	n=0.52
	k=3	n=0.79	n=0.39
	k=4	n=0.51	n=0.14
	k=5	n=0.25	n=0.03

In that region of the β-spectrum in which the measurements of e/m_0 were carried out by the writers, most of the β-particles are emitted by RaB, whereas the data given above show the ratio of the number of heavy electrons to that of RaC electrons. If this be taken into account, then

		if a=1	if a=2
for	k=2	n=0.33	n=0.17
	k=3	n=0.26	n=0.13
	k=4	n=0.17	n=0.047
	k=5	n=0.084	n=0.01

The measurements were carried out with the electric field of 3000 V, and a varying magnitude of the magnetic field.

Similar measurements were carried out with β-particles of RaE, analogous results being obtained.

Experiment shows that if during β-disintegration heavy electrons, having masses from 2 to 5 m_0, are also emitted in addition to ordinary electrons, then the number of those heavy electrons is at least 200 times smaller than that of the ordinary ones. This ratio is many times smaller than the ratios calculated from equation (1) and given in the second table. Apparently, this would mean that electrons with masses of 2, 3, 4 and 5 m_0 do not exist in nature, since otherwise it is difficult to understand why they should not be emitted in β-disintegration, i. e. in a process in which the magnitude of the mass of the emitted particles cannot play any important part.

In conclusion it should be noted that the above results refute Jauncey's statement as to the existence of a continuous spectrum of masses of the electrons emitted by RaE.

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PHYSICS

THE SPECTRUM OF RaC POSITRONS

By A. ALICHANOV and G. LATYSHEV

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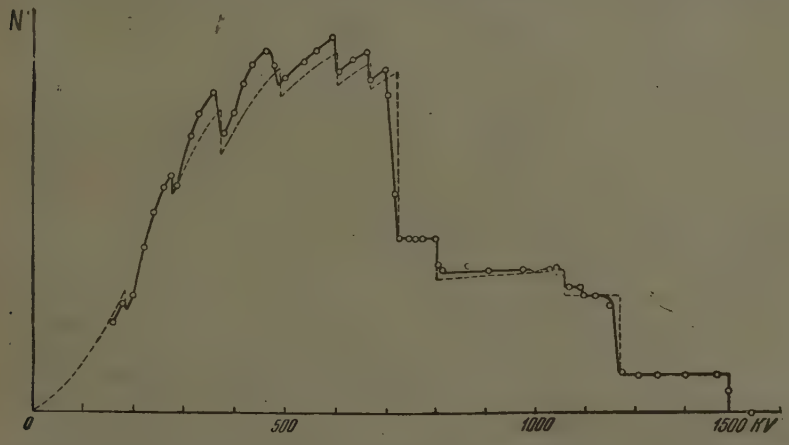
The investigation of spectra of RaC and Th ($C+C''$) positrons carried out in our laboratory for a number of years invariably led us to the conclusion that the form of these spectra cannot be explained by the internal conversion



of γ-rays alone. However, the repeated investigations of the spectrum of Th ($C+C''$) positrons, made by Alichanov and V. Dželepov (1) under improved conditions and with a more powerful source, have shown that the discrepancy between the theoretical form of the positron spectrum, calculated according to Jaeger and Hulme (2), and the experimental form is accounted for by a great number of so far unknown γ-lines which have formed the so-called «superfluous positrons». We decided therefore to repeat the investigation of the spectrum of RaC positrons which has also shown a great number of those superfluous positrons.

In order to detect even the faint lines it was necessary to improve the method of investigation so that a more powerful source could be used. However this is limited by the fact that in a Geiger-Müller counter arise a great number of discharges from γ-rays of the source; and besides, owing to the low resolving power of the counters themselves, not all the positrons passing through

both counters are recorded. The system has no time to recover and the next electron (positron) passing through it is unable to produce a new discharge.

In order to avoid this hindrance M. Kozodaev and G. Latyshev have designed a peculiar scheme which facilitates the use of counters with very low resistances. This scheme is described in a paper by M. Kozodaev and G. Latyshev [3] and we shall not dwell on it here. On the basis of that scheme a special amplifier was constructed to separate and record coincidences. This amplifier was used in the present experiments.

The investigation of the spectrum was carried out by a magnetic spectrograph in which the positrons were focussed by Danisz's method and recorded by the method of coincidence in two counters.

The results of the experiment are shown in the figure. The eleven abrupt drops of the curve correspond to the eleven γ-lines, the energies of which are given in the table.

No	Position of the drop in the curve in ekV	Energy of γ-lines in ekV	Intensity of bands in relation to γ-lines in 2200 ekV
1	190	1 210	0.27
2	270	1 290	0.44
3	370	1 390	1.19
4	500	1 520	0.71
5	600	1 620	0.54
6	670	1 690	0.40
7	730	1 750	2.42
8	800	1 820	0.41
9	1 070	2 090	0.37
10	1 180	2 200	1.00
11	1 400	2 420	0.50

Our previous measurements enabled us to detect only 4 drops, i. e. the γ-lines of 1 390, 1 750, 2 900 and 2 400 ekV.

Ellis has also observed by the method of internal conversion on the electron shell a great number of newly detected γ-lines, however, he has undervalued their intensities. The intensities of these lines are shown in the fourth column in the table and are calculated on the basis of the theoretical curve showing the relation between the internal conversion coefficient and the energy of γ-rays, as calculated by Jaeger and Hulme. It was also assumed that all the lines are of quadripole origin.

The scheme of levels of the energy of RaC suggested by Ellis can comprise all these γ-lines, the transitions corresponding to these lines being quadripole.

The results obtained have shown that the study of the spectrum of positrons emitted by radioactive substances is the most suitable method for the study of the spectrum of γ-rays with energies higher than $2\,mc^2$.

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25. VI. 1938.

REFERENCES

[1] Alichanov a. V. Dželepov, C. R. Acad. Sci. URSS, XX, № 1 (1938). Jaeger a. Hulme, Proc. Roy. Soc., 148, 865 (1935). [3] Kozodaev a. Latyshev, C. R. Acad. Sci. URSS, XX, № 1 (1938).

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PHYSICS

WAVE LENGTH MEASUREMENTS OF DEUTERIUM AND HYDROGEN BALMER LINES. I

By Vs. ČERNIAJEV

(Communicated by S. I. Wawilow, Member of the Academy, 28.VI.1938)

No exact measurements have been made so far except for the four first members of the Balmer series of deuterium (1). In the present work patterns of the first ten lines of the Balmer series were obtained by means of a large diffraction grating of 640 cm radius of curvature, with the total number of 180 000 lines, 30 000 rulings per inch. The patterns were obtained in the first order of the grating with a tube fed from a transformer and filled either with a mixture of vapours of heavy and light water, or with helium, in which traces of hydrogen and deuterium were present (2). On the same plates the spectrum of an iron arc was also superimposed, thus giving standard wave lengths, by means of which, using an interpolation formula with a quadratic member, the measured wave lengths of hydrogen and deuterium were calculated.

Making use of the theoretical values for the intensities and the mutual positions of the fine structure components of Balmer lines, Penney (3) derived a formula for the wave numbers of the centres of gravity of these lines.

Substituting in Penney's formula the n values of the upper levels and the value of α^2, it proved possible to determine the R_H and R_D values from the measurements of ν of each line.

In the table are given average values of the wave lengths measured on several plates, corresponding wave numbers and Rydberg's constants calculated by means of Penney's formula. The data for H_α and D_α lines are not included in the table, accurate measurements of these lines being prevented by a wide doublet structure, and also owing to the absence in their vicinity of any convenient iron standards. The fine structure of these lines was studied in a previous work (4) for the exact determinations of Rydberg's constants.

The values of Rydberg's constants as obtained from the measurements of the fine structure of D_α and H_α lines are (4):

$$R_D = (109\ 707.56_7 \pm 0.030)\ \text{cm}^{-1},$$

$$R_H = (109\ 677.75_3 \pm 0.030)\ \text{cm}^{-1}.$$

Thus, the R values, determined from the measurements of the centres of gravity of the higher members, differ essentially from those obtained from fine structure data. This may result first from an insufficient accuracy of the

measurements of diffuse H and D lines. An error of 0.005 Å (which is not impossible), if made near 3800 Å, would lead to an error of 0.18 cm^{-1} in the R value. Then, in measuring diffuse lines it is difficult to say whether the centre of gravity, or, merely, the middle of a wide line, is being measured. Deviations from the theoretical values of the intensity distributions among the fine structure components of the lines are also possible, the intensities being greatly affected by discharge conditions, more particularly, by the shape of the discharge tube (5). The agreement of the component intensities with theoretical ones for the lower members (4) had been found with a tube of another shape. Moreover, if the width of the components is a finite one, the maximum of intensity does not necessarily coincide with the calculated centre of gravity (6).

n	Line	λ air, Å	ν vac., cm^{-1}	R_D cm^{-1}	R_H, cm^{-1}
12	D_{10}	3749.153	26665.140	109706.877	—
12	H_{10}	3750.145	26658.086	—	109677.855
11	D_9	3769.623	26520.345	109706.963	—
11	H_9	3770.640	26513.192	—	109677.373
10	D_8	3796.871	26330.026	109707.306	—
10	H_8	3797.902	26322.880	—	109677.531
9	D_7	3834.352	26072.656	109707.177	—
9	H_7	3835.394	26065.573	—	109677.374
8	D_6	3887.992	25712.958	109707.471	—
8	H_6	3889.048	25705.976	—	109677.681
7	D_5 (ε)	3968.990	25188.227	109707.560	—
7	H_5 (ε)	3970.071	25181.368	—	109677.686
6	D_4 (δ)	4100.621	24379.694	109707.444	—
6	H_4 (δ)	4101.731	24373.097	—	109677.758
5	D_3 (γ)	4339.277	23038.858	109707.638	—
5	H_3 (γ)	4340.458	23032.590	—	109677.791
4	D_2 (β)	4859.992	20570.438	109707.729	—
4	H_2 (β)	4861.322	20564.809	—	109677.708
			Average	109707,352	109677.640

The table shows that the R values for deuterium are, probably, increased when passing from the higher to the lower series members taken for the calculation of R. Though the exactness of R determinations decreases towards the higher members, the probability of such a monotonous change in the calculated R values still exists. This, possibly, points to a systematic increase of deviations from theoretical component intensities when passing to higher members. The lack of accuracy of measurements for hydrogen lines, which are more diffuse than those of deuterium, is greater, thus, this effect is less pronounced for these lines.

The average R values of the table are smaller than those obtained from fine structure data, this difference for R_H amounting to 0.113 cm^{-1} and for R_D even to 0.215 cm^{-1}. These differences lie near the limits of error of measurements for R as determined from separate lines, however, we have averaged here the R values for nine different lines. It is therefore quite possible that these differences are true. The different deflection for both isotopes may point to unequal distinctions of the intensities of fine structure components from the theoretical ones and to greater distinctions for deuterium.

The lower values of R calculated from higher members may also result

from the influence of the Stark-effect due to the electrical fields. The first order Stark-effect results in symmetrical splitting of the lines. With the second order effect (7) all terms are shifted towards the lower energy values, these shifts being rapidly increased with the increase of n. This must lead to the decrease of R values calculated from the higher members of the series. These considerations hold true for strong fields, but, beginning with the 6th level (8), non-homogeneous tube fields, under usual experimental conditions, suffice to establish some kind of statistical equilibrium between different fine structure sublevels. Thus, for high levels rules derived from the assumption of strong fields are practically always applicable.

Notwithstanding low tube fields, the strong dependence of the second order effect on n may give an appreciable effect for high members. The above considerations can, however, not account for the difference in the deflections of the average R_H and R_D values from fine structure data. Thus, there still remains some possibility that the intensity distributions among the fine structure components of the higher members (or, rather, the Stark-components) of the hydrogen and deuterium series are different.

These differences cannot be regarded as a proof of the isotope shift of hydrogen lines being essentially influenced by the size of the nucleus, which was assumed to play a part in the isotope shift of more complex atoms (9). All Balmer lines having components of the same serial nature must have the same sensitivity for isotope shift (such as takes place for complex atoms). On the other hand, the influence of nucleus size must decrease with the increase of quantum number n, since the density of the electron cloud in the nucleus decreases with an increasing n. Thus, the R_D value as calculated from the measurements of high members would be closer to the value obtained from the elementary theory of «Mitbewegung des Kernes», than if calculated from the measurements of low members. If, however, using formulae of the elementary theory, we express R_D in terms of R_∞, R_H and proton-deuton mass ratio, and take Birge's values for R_∞ and R_H (10), and Aston's values for masses (11), the calculated value of R_D coincides, within the limits of error, with the value obtained by the present author from the fine structure measurements of D_α, for which the influence of nucleus size ought to have been the greatest.

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REFERENCES

[1] H. C. Urey, F. C. Brickwedde a. G. M. Murphy, Phys. Rev., **39**, 164, 536; **40**, 1 (1932); see also H. C. Urey a. G. K. Teal, Rev. Mod. Phys., **7**, 34 (1935). [2] Vs. Černiajev, C. R. Acad. Sci. URSS, XIX, 245 (1938). [3] V. G. Penney, Phil. Mag., **9**, 661 (1930). [4] Vs. Černiajev, C. R. Acad. Sci. URSS, XX, № 5 (1938). [5] G. E. Harrison, Nature, **119**, 393 (1927). [6] R. C. Williams a. R. C. Gibbs, Phys. Rev., **45**, 475, 491 (1934). [7] H. Bethe, Hb. d. Phys., XXIV, Erster Teil, § 30, 31. [8] H. Bethe, loc. cit., § 43. [9] J. H. Bartlett, Nature, **128**, 408 (1931); G. Racah, ZS. f. Phys., **71**, 431 (1931); Nature, **129**, 723 (1932); J. E. Rosenthal a. G. Breit, Phys. Rev., **41**, 459 (1932); G. Breit, Phys. Rev., **42**, 348 (1932); **46**, 319 (1934); E. C. Kemble a. R. D. Present, Phys. Rev., **44**, 1031 (1933). [10] T. R. Birge, Rev. Mod. Phys., **1** (1929). [11] F. W. Aston, Nature, **137**, 357 (1936).

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PHYSICS

THE INFLUENCE OF NOBLE GASES ON THE INTENSITY OF LINES OF HYDROGEN AND DEUTERIUM BALMER SERIES. II

By Vs. ČERNIAJEV

(Communicated by S. I. Wawilow, Member of the Academy, 28.VI.1938)

In an earlier paper by the same author (1) the results of measurements of the intensity ratios of Balmer lines of deuterium to respective lines of hydrogen were given, the discharge taking place either in water vapour, or in noble gases (He, Ne, Ar) with traces of hydrogen and deuterium present in the discharge. The measurements of some new plates yielded further confirmation of the earlier results as well as more reliable data at to the intensity ratios of δ-lines for the discharge in water vapour and in He. The average values of the intensity ratios of D lines to the corresponding H lines for all plates measured are given in the table.

Tube content	H_2O	He	Ne	Ar
$\frac{D_\alpha}{H_\alpha}$	1.14	1.48	1.19	1.20
$\frac{D_\beta}{H_\beta}$	1.14	1.60	1.11	1.19
$\frac{D_\gamma}{H_\gamma}$	1.14	1.80	1.12	1.15
$\frac{D_\delta}{H_\delta}$	1.14	1.99	—	—

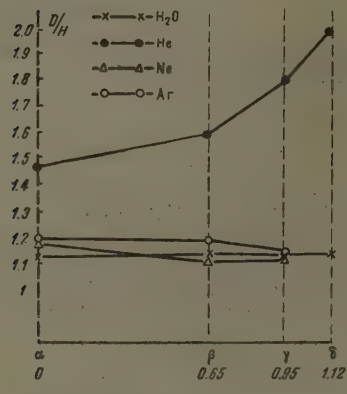
A graph of these results is shown in the figure, where the ratios of line intensity are plotted against the corresponding upper levels, the distances between the latter being taken in volts beginning with the three-quantum state.

In the water vapour discharge the $\frac{D}{H}$ ratios are wholly independent of the number of the series member. In the He discharge a general enhancement of D-lines, as well as a marked increase of $\frac{D}{H}$ ratios towards higher members, is to be seen. In Ne one can see a decrease of the relative intensity of the D-line in the H_β, D_β pair, as compared with the H_α, D_α pair. The drop of the curve from β to γ in the Ar discharge is scarcely real, because under the conditions of the experiment, the H and D spectra in the Ar discharge were in most cases weak, the γ-lines lying on the lower part of the density curves, plotted from microphotometric records. When the γ-lines were strong enough [see (1), plates 6 b and 7 b], no such drop was observed.

The enhancement of the higher series members of deuterium in He discharge was referred to (1) a more intense recombination spectrum of D than that of H under the conditions of the experiment. Indeed, when a great number of ions is known to be present in the discharge, the higher members of the

series are very prominent (2). But the possibility of this recombination spectrum being due to collisions resulting in a simultaneous excitation of the two electrons of the D_2 molecule, was based on a misconception, since when the position of the potential curve of such an excited molecule was calculated, the dissociation energy of the D_2 (or H_2) molecule was not taken into account. This energy being added, the asymptotic value of the potential curve of the excited molecule is found to lie about 24.7 V above the ground state of the unexcited molecule. This value is much higher than the metastable potential of He.

An explanation of the effects observed in the He discharge should probably be sought for in the difference between the masses of the two isotopic molecules. As had been pointed out (1), even in case of a large distance between the metastable helium potential and the limit of convergence of the vibrational levels of the ground state of the D_2^+ (H_2^+) ion, the ionization of D_2 (H_2) molecules, through collisions with the metastable He atoms, may occur. Some earlier works (3) show that in case of collisions of metastable atoms (A) with others (B), the ionization potential V_i ot the latter being lower than the metastable potential V_{met} of the former, an ionization of B atom occurs even when the energy difference $V_{met}-V_i$ is rather great. The excess of energy may in our case be distributed among the kinetic energy of the liberated electron and the energy of dissociation of D_2^+ (H_2^+) ion. The second metastable potential of the helium atom (20.55 V), also lying in the continuum region of D_2^+ (H_2^+) ion, may play some part in collisions, too. However, this part is probably less than that of the first metastable level, since the second one is farther from the limits of vibration levels of D_2^+ (H_2^+)ions. Collisions with He^+ ions (ionization potential 24.47 V), namely processes of neutralization of He^+ ions by an electron captured from H_2 or D_2, may also play some part (3).



If the effectiveness of collisions resulting in the dissociation and ionization of H_2 and D_2 molecules were dependent only on the distance between He levels and the limit of convergence of the ground states of H_2^+ or D_2^+ ions, then the slight difference between the limits of H_2^+ and D_2^+ (0.07—0.08 V), with great distances to He levels, would result in no essential variations of the effectiveness of collisions with H_2 and D_2. Yet, as was pointed out (1), the difference between the reduced masses of colliding particles, if appreciable, may also intervene here. There are known cases of a marked effect of «mass resonance» when atomic collisions take place (4). This resonance may in the present case result in a noticeable enhancement of the deuterium recombination spectrum as compared with that of hydrogen.

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REFERENCES

[1] Vs. Černiajev, C. R. Acad. Sci. URSS, XIX, 245 (1938). [2] G. Herzberg, Ann. d. Phys., **84**, 553—565 (1927); G. W. Fox a. C. H. Bachmann, Phys. Rev., **50**, 939 (1936); F. L. Mohler, Phys. Rev., **31**, 187 (1927); Rev. Mod. Phys., **1**, 216 (1929). [3] W. de Groot u. F. M. Penning, Hb. der Phys., XXIII, 1 Teil (1933). [4] O. Beec a. J. C. Mouzon, Phys. Rev., **38**, 967 (1931); Ann. d. Phys., **11**, 737, 858 (1931); Proc. Nat. Acad., **18**, 311 (1932); ZS. f. Phys., **76**, 799 (1932).

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PHYSICS

ON THE EFFECT OF MAGNETIC FIELD UPON MERCURY DISCHARGE RADIATION. II

By V. FABRIKANT and G. ROKHLIN

(Communicated by S. I. Wawilow, Member of the Academy, 28.VI.1938)

The object of the present work is to investigate by optical methods the mechanism of the discharge in a longitudinal magnetic field. The possibility of such an application of optical methods has already been schemed in our previous work (1). As compared with the method of electric probes, the optical method has the advantage of being applicable irrespective of the presence of the magnetic field. Its defect consists in that it permits of investigating the behaviour of relatively fast electrons alone (with energies exceeding the excitation potential). It will therefore be advantageous to compare the results of optical measurements with those obtained by the probe method.

1. The measurements were carried out with a mercury discharge tube of a special design that had already been applied for a series of experiments made in our laboratory (2).

This tube is provided with a liquid cathode; the vapour pressure therein is being maintained by means of a water bath. The tube is thus designed as to eliminate the fast electrons flying out of the cathode part of the discharge.

The magnetic field was created by means of two solenoids put around the straight part of the tube. The dimensions of the solenoids were as follows: the radius of the inside aperture—36 mm; the outside radius—62 mm; length—62 mm. The solenoids were placed 22 mm apart. The observations were made within the interspace between the solenoids.

All the measurements were made at a temperature of the water bath equal to 18° C and with a 0.5 A current. The conditions thus selected made it possible to reduce to the minimum the rôle of secondary processes. The intensity distribution across the discharge section in the light of separate spectral lines — 5461 Å ($7^3S_1 \rightarrow 6^3P_2$), 4358 Å ($7^3S_1 \rightarrow 6^3P_1$), 5791 Å ($6^1D_2 \rightarrow 6^1P_1$), 3906 Å ($8^1D_2 \rightarrow 6^1P_1$), 3704 Å ($9^1D_2 \rightarrow 6^1P_1$), 4916 Å ($8^1S_0 \rightarrow 6^1P_1$), 4108 Å ($9^1S_0 \rightarrow 6^1P_1$) — was measured.

Intensity measurements were effected by the photographic method, a diminished horizontal image of the lamp being focussed upon the slit of the spectrograph.

2. The solid curves in Fig. 1 and 3 and the dotted one in Fig. 2 show the results of the measurements.

The curves in Fig. 1 give a clear idea of the contracting of the discharge in the light of the green line 5461 Å with the increase of magnetic field.

Along the axis of abscissae are plotted the displacements in segments of the tube radius, along the axis of ordinates—the brightnesses.

The curves of Fig. 3 show the influence of a magnetic field on the maximum brightness (in the centre of the tube) for the lines 5791 Å, 3906 Å and 3704 Å. The lines of the singlet sharp series 4916 Å and 4108 Å exhibit a more sloping drop of intensities.

3. The results of optical measurements quoted above permit of drawing quantitative conclusions as to the influence of a magnetic field both upon the electron distribution across the section of the discharge tube and the mean electron energy. It should be noted once more that this refers only to the electrons with energies exceeding the lower excitation potential.

The coincidence of the curves of distribution for lines 5791 Å and 3704 Å with different excitation potentials points to the stability of electron energy within the limits of the contracted discharge column.

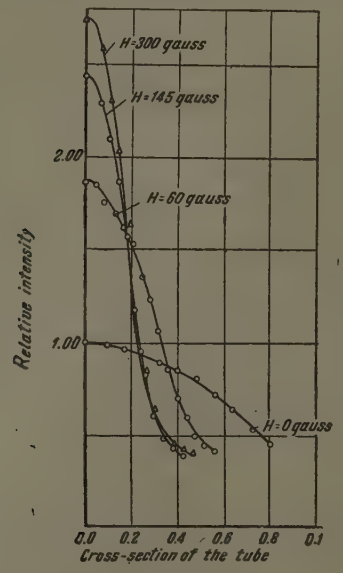

Fig. 1.

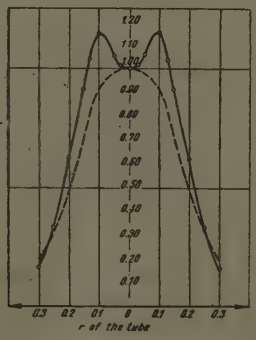

Fig. 2.

The distribution of the volume density of the radiation emitted by the discharge should therefore coincide with that of the volume density of electrons.

The distribution of the volume density of radiation can, certainly, not be derived directly from the dotted curve in Fig. 2; yet, after due treatment, such a distribution can be calculated from that curve. As a matter of fact, when a discharge column is photographed, the brightness of each point represents the result of the action of an entire level. The transition from these integral brightnesses to volume densities of the radiation is reduced to the solution of an integral equation of the Abel type (3).

The solid curve in Fig. 2 represents the result of such a recalculation. According to the abovesaid, this curve coincides with the distribution of electron density across the section of the discharge tube. Thus, the maximum density of electrons is being observed at some distance from the axis*. A qualitative interpretation of such a distribution may be easily obtained using Störmer's theory (4). The electrons are accumulated between two forbidden zones.

* We dispose of unofficial information from G. V. Spivak who pointed to the same effect as having been shortly observed by him and E. M. Reichrudel using the probe method.

As it is stated that the mean energy of electrons across the discharge action is stable, there is no difficulty to obtain from the curves in Fig. 3 the run of the mean energy of electrons depending on the strength of the magnetic field.

In our previous work we have already noted that in the line 5770—91 Å, characterized by a higher excitation potential than the visible triplet, there may be observed a slight drop of intensity with the increase of magnetic field. The curves in Fig. 3 show quite distinctly that the intensities of the lines having high excitation potentials are decreasing with the growth of magnetic field. Besides, the drop is the sharper the higher the excitation potential (5791 Å — 8.8 V; 3906 Å — 9.7 V; 3704 Å — 10 V).

In so far as under the investigated conditions of the discharge the secondary effects could play no appreciable part, the curves in Fig. 3 represent the direct change in the number of exciting collisions with the growth of magnetic field. One may as well assume that excitation mostly proceeds from the normal state.

It is known that the number of exciting collisions depends both on concentrations and energies of electrons and on the excitation function ([5]).

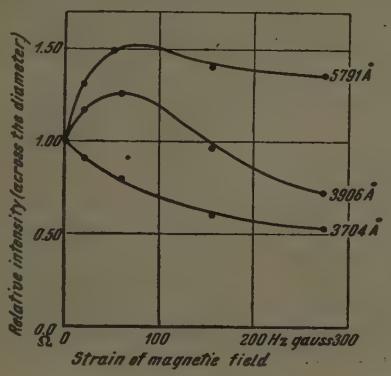

Fig. 3.

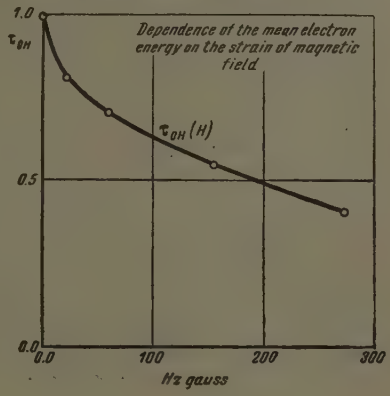

Fig. 4.

The run of the curves in Fig. 3 may be accounted for only by the fact that with the increase of the strength of a magnetic field the concentration of electrons is growing too, while their mean energy is decreasing.

Taking the relation of ordinates of two curves, we obtain values that are depending only on the electron energy and on the form of the function of excitation ([5]). The lines 5791, 3906 and 3704 Å possess typical singlet excitation functions which materially simplifies the quantitative calculation of the run of electron energy from the curves in Fig. 3. In Fig. 4 is shown the relative run of the mean energy of electrons with the growth of magnetic field thus obtained. The initial energy of electrons in the absence of magnetic field equals 3.5 V ([6]).

We see that the energy of electrons is monotonously decreasing with the growth of magnetic field. It is noteworthy that a less abrupt drop of intensities in the lines 4916 and 4108 Å may be fairly well accounted for by the form of their respective excitation functions, being at the same time significative of the correctness of the assumption that led to the determination of the drop of electron energy. The lines 4916 and 4108 Å display typical triplet excitation functions (the second maximum of the excitation function in the discharge being unessential) ([7]) in spite of their originating in singlet levels.

In connection with the drop of electron energy found by us, we want to briefly dwell upon the results obtained by Slutzkin ([8]). Subscribing to the criticism the methods of this work ([9]) underwent to, it should be added that the phenomenon revealed in this work is virtually no influence of magnetic field upon the electron energy in the discharge plasm, but merely a decrease of electron energy accompanying the diffusion in gas.

The rôle of the magnetic field was reduced in those experiments to a simple production of a filiform source of electrons. From this point they are analogous to the experiments with incadescent filament serving as source of electrons.

Results

1. The contracting of a mercury discharge of low pressure by a longitudinal magnetic field was subjected to quantitative investigation.

2. The distribution of electrons over the discharge cross-section within the longitudinal magnetic field was determined with the help of the optical method; a maximum concentration of electrons was found to lie at some distance from the axis of the discharge.

3. The decrease of the mean energy of electrons in the discharge plasm produced by a longitudinal magnetic field was established and studied by means of optical methods.

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REFERENCES

[1] V. Fabrikant a. G. Rokhlin, C. R. Acad. Sci. URSS, XIX, № 5 (1938). [2] W. Fabrikant, F. Butaewa u. I. Zirg, Sow. Phys., **11**, 576 (1937). [3] H. Hörmann, ZS. f. Phys., **97**, 539, (1935); W. Elenbaas, Physica, **2**, 757 (1935). [4] C. Störmer, Über die Probleme des Polarlichtes, Erg. d. kosm. Physik (1931). [5] V. Fabrikant, C. R. Acad. Sci. URSS, XV, № 8, 451 (1937). [6] B. Klarfeld, Techn. Phys. URSS, **4**, 44 (1937). [7] W. Schaffernicht, ZS. f. Phys., **62**, 106 (1930). [8] A. Slutzkin, Sow. Phys., **12**, 215 (1937). [9] Г. Спивак и Э. Рейхрудель, Ж. Э. Т. Ф., 8, 319 (1938).

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PHYSICS

EXCITATION OF ATOMS IN A RARE GASES DISCHARGE

By V. FABRIKANT and K. PANEVKIN

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1. Kopfermann and Ladenburg (1), as well as Krebs (2) and Schön (3), have investigated by means of optical procedures the influence of current intensity upon the concentration of excited atoms in rare gase discharges.

Kopfermann and Ladenburg have found that the concentration of excited atoms at 3^1P_1, $3^3P_{0,1,2}$ levels of neon are increasing with current intensity until they have reached certain maximum values, after which they become reduced again in spite of the continuing growth of the current intensity.

Working with high densities, Krebs observed only the second part of the curve for neon, i. e. that of the drop of concentration with increase of the current. Schön, on the other hand, has observed for argon the first part of the curve—that of concentration increase and saturation. The experimental results of all the three investigations agree with each other*.

Yet, being subjected to a more detailed analysis, the theoretical interpretations given by Kopfermann and Ladenburg to their experimental data (interpretations that, by the way, have received general recognition) prove more than questionable.

The present work can be divided into two parts: the first contains a criticism of Kopfermann-Ladenburg's theory, while in the second one a new interpretation of the experimental results obtained by Kopfermann and Ladenburg, Krebs and Schön is suggested. At the same time an attempt is made to clear up the rôle of individual factors in the whole mechanism of atomic excitation in a rare gas discharge.

2. Kopfermann and Ladenburg believe that the saturated concentration they have observed in excited atoms is to be accounted for by the setting in of an equilibrium between electronic collisions of the first and second kind. Hence, maximum concentrations should satisfy Boltzmann's formula, having in the exponent the electronic temperature. Kopfermann and Ladenburg compute these temperatures from the concentration of excited atoms, using Boltzmann's formula, and compare the values obtained with the data of probe measurements (4). As a matter of fact, the values obtained are rather congruent, which convinces these authors of the correctness of their judgment.

However, the comparison made by Kopfermann and Ladenburg is lacking validity, since the discharge conditions under which their experiments were conducted do sharply differ from those prevailing in the probe measurements effected by Seeliger and Hirschert. The diameter of Kopfermann and Ladenburg's discharge tube equaled 10 mm, while that of Seeliger and

* After the present paper had been written, appeared an article by Kruse (13) dealing with analogous measure ments in helium. The results of this work agree in the main with those of his earlier works.

Hirschert was as large as 40 mm. Thus, the potential gradient in a positive column in Kopfermann and Ladenburg's experiments considerably exceeds that of Seeliger and Hirschert (>3.5 V/cm and≈2 V/cm respectively). Under these conditions it may be assumed as usual ([5]) that electronic temperatures are proportional to the gradients (at a constant pressure). Consequently, Kopfermann and Ladenburg used for their comparison rather understated values of electronic temperatures. From the data by Seeliger and Hirschert and from the ratio of gradients we obtain for the discharge tube an electronic temperature equal to 4.6 V, which is nearly twice that calculated by Kopfermann and Ladenburg from the results of optical measurements (2.6 V). Similar calculations may be done for other currents under the conditions of Krebs and Schön's experiments, making use of the results of probe measurements in neon at high current intensities ([6]), and in argon too ([7]).

If we take the electronic temperatures determined in the manner described above, and substitute them into Boltzmann's formula, we then obtain concentrations surpassing the values experimentally observed by several orders of magnitude. Table 1 contains the results of such calculations. As should be expected, the discrepancy proves to be smaller at large currents, but whatever be the conditions, experimental data prove to be considerably lower than those calculated after Boltzmann.

Table 1

Comparison of Experimental Data for Concentration with Those Calculated after Boltzmann

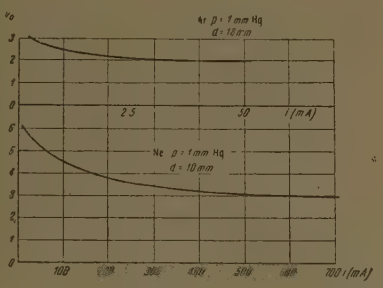
Observed	In the maximum of concentration		At the maximum current intensity	
	Exper.	Boltzmann	Exper.	Boltzmann
Kopfermann and Ladenburg	$2 \cdot 10^{12}$	$8 \cdot 10^{14}$	1.10^{12}	$5 \cdot 10^{13}$
Krebs	$5 \cdot 10^{10}$	$3 \cdot 10^{13}$	$3 \cdot 10^{10}$	$3 \cdot 10^{10}$
Schön	$8 \cdot 10^{11}$	$6 \cdot 10^{13}$	$8 \cdot 10^{11}$	$1.5 \cdot 10^{13}$

From Table 1 may be seen that the conditions observed in all the experiments are far from being those of temperature equilibrium. Thus the saturated concentration found to vary with the current intensity cannot be referred to the setting in of equilibrium between collisions of the first and second kind.

3. The correctness of the conclusion made at the close of § 2 may be proved quite independently in another way. The saturation of concentrations may be accounted for by the setting in of an equilibrium between the collisions of the first and second kind in that case only, when in the maximum of concentrations the destruction of the excited atoms proceeds almost totally through collisions of the second kind with electrons. If that is the case, the probability of the second kind collisions with electrons should considerably exceed the probabilities of other destruction processes. For sake of clearness it may be arbitrarily assumed that equilibrium is established, if the probability of second kind collisions exceeds the probabilities of destruction of excited atoms through diffusion towards the walls by an order of magnitude ($B_a > 50$). [Under the probability of second kind collisions we understand the mean number of second kind collisions suffered by an atom per second ([8])].

Making use of the results of probe measurements and of the values of the probabilities thus determined, one succeeds in determining the effective cross section of excited atoms for the second kind collisions Q_{21} (0) ([8]). In the third column of Table 2 are given Q_{21} (0) calculated by this method.

On the other hand, the effective cross-sections $Q_{21}(0)$ may be determined from the absolute concentrations of excited atoms with small currents. For this purpose it is sufficient to make a calculation inverse to that previously made by one of us for the concentrations of excited atoms ([8]). As a matter of fact, the calculation is based on that the ratio of the concentration of excited atoms to their life time is equal to the number of exciting collisions. From the number of the latter and the results of probe measurements the effective cross-section of atoms for the collisions of the first kind $Q_{12}(Vm)$ may be readily calculated (within the maximum of the excitation function). In so far as those atomic constants present independent interest, they are given in column 1 of Table 2. From $Q_{12}(Vm)$ it is easy to determine $Q_{21}(0)$ ([9]) having recourse to Klein-Rosseland's relation. In the second column are given $Q_{21}(0)$, determined in this way. These values are in good agreement with the values of $Q_{21}(0)$ for the atoms of metals ([9]) and practically approach to the real values of the atomic constants quoted above.



We thus become aware that $Q_{21}(0)$, determined from saturation, exceeds the real values of $Q_{21}(0)$ for three orders of magnitude. The discrepancy observed points to the fact that saturation must be due to some other cause rather than to collisions of the second kind. The probability of the occurrence of the latter is too insignificant for a saturation of concentrations to be called forth under the conditions investigated. The discrepancy obtained makes it obvious that the effect of the real cause responsible for saturation is equivalent to the increase of the probability of second kind collisions for three orders of magnitude. We thus again arrive at the conclusion that the saturation of concentrations under the conditions of Kopfermann and Ladenburg's as well as Schön's experiments cannot be caused by the establishment of equilibrium between collisions of the first and second kinds.

Table 2

Effective Cross-Sections of Atoms of Neon and Argon

Gas	$Q_{12}(Vm)_{cm^2}$	$Q_{21}(0)_{cm^2}$	
	From conc. with small currents	From conc. with small currents	From saturations
Ne (Kopfermann a. Ladenburg) . .	$2.5 \cdot 10^{-18}$	$4 \cdot 10^{-18}$	$9 \cdot 10^{-15}$
Ar (Schön) . . .	$1.5 \cdot 10^{-17}$	$2 \cdot 10^{-17}$	$1.4 \cdot 10^{-14}$

4. Now what is then responsible for the course of concentrations of the excited atoms proceeding with the current intensity as observed by Kopfermann and Ladenburg, Krebs and Schön?

This question can be easily answered, if we take into account that the atoms of rare gases possess a high excitation potential. Owing to high excitation potentials, the number of collisions of the first kind between electrons and atoms of rare gases is strongly dependent on electronic temperature. An insignificant change in electronic temperature is sufficient to produce a sharp decrease of the number of exciting collisions. Yet it is known that the growth of current intensity calls forth a marked reduction of the electronic temperature in the discharges which take place in rare gases ([4]). Thus, the current increase is, on one hand, accompanied by the increase of electron concentrations, on the other hand—by the fall of electronic temperature.

The superimposition of both factors gives a maximum for the number of exciting collisions at a given current intensity. With a constant life time of the excited atoms the course of their concentrations should simply coincide with that of the number of exciting collisions, the current intensity being changed. Such an interpretation of the course of concentration experimentally observed is diametrically opposite to that suggested by Kopfermann and Ladenburg, their interpretation being based essentially upon the shortening of the life time of excited atoms at the expense of collisions of the second kind.

Using the relations mentioned in § 3, one may estimate what should be the drop of electronic temperature to produce the course of concentrations described in the articles quoted above. In the figure are shown the results of such a calculation made on the basis of data by Kopfermann, Ladenburg and Schön; it may be seen that quite an insignificant drop of electronic temperature with the current suffices to account for all the effects observed. An analogous drop of electronic temperature with the current has been observed directly in discharges in rare gases. The view we are evolving in the present article provides explanation that may be extended to the whole run of the experimental curves. In particular, to explain the drop of concentrations in the region of large currents it is not necessary to have recourse to a new process, say, that of destruction of excited atoms through collision with each other (as Kopfermann and Ladenburg do). The differences in the behaviour of separate levels are accounted for by different forms of excitation functions.

At first sight it may appear that the rise of concentrations at the upper levels 3^1S_0, 4^1P_1, 3^3S_1, $4^3P_{0,1,2}$, $3^3D_{1,2,3}$, 3^1D_2 etc., as found in control experiments ([10]), is conflicting with the interpretation just suggested, in so far these levels possess higher excitation potentials. This is, however, not the case. As shown by Dorgelo ([11]) the excitation of those levels proceeds mostly stepwise, and if we calculate the course of the number of acts of the stepwise excitation with current, an uninterrupted increase is obtained. It may be put down to the fact that the excitation of upper levels begins with the levels 3^1P_1, $3^3P_{0,1,2}$. The excitation potential in the case of such processes is of the order 2—3 V, and the reduction of electronic temperature does practically not tell upon the number of excitation acts.

From all stated above may be seen that the conditions of Kopfermann and Ladenburg's experiments do anything but approach temperature equilibrium: the effect called forth by negative absorption should be considerably smaller than it was assumed by them. This question has elsewhere been treated by one of us ([12]).

Conclusions. 1. The saturation of the concentrations of excited atoms observed by Kopfermann and Ladenburg, Krebs and Schön, cannot be accounted for by the equilibrium established between the collisions of the first and second kind. 2. The main cause responsible for the course of the concentrations of excited atoms as observed by the above authors is the drop of electronic temperature with current.

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Received
29.VI.1938

REFERENCES

[1] H. Kopfermann, R. Ladenburg, ZS. f. Phys., **65**, 167 (1930). [2] K. Krebs, ZS. f. Phys., **101**, 604 (1936). [3] K. Schön, Ann. d. Phys., **28**, 649 (1937). [4] R. Seeliger, R. Hirschert, Ann. d. Phys., **11**, 817 (1931). [5] J. S. Townsend, Jrnl. of Frankl. Inst., **200**, 563 (1925). [6] M. J. Druyvesteyn, ZS. f. Phys., **81**, 571 (1933). [7] A. Lompe, R. Seeliger, Ann. d. Phys., **15**, 300 (1932). [8] V. Fabrikant, C. R. Acad. Sci. URSS, XIX, 389 (1938). [9] V. Fabrikant, C. R. Acad. Sci. URSS, XVII, 245 (1937). [10] R. Ladenburg, S. Levy, ZS. f. Phys., **65**, 189 (1930). [11] N. B. Dorgelo, Physica, **5**, 90 (1930). [12] V. Fabrikant, Bull. Acad. Sci. URSS (1938). [13] M. Kruse, ZS. f. Phys., **109**, 312 (1938).

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PHYSICS

INVESTIGATION OF THE PHOSPHORESCENCE MECHANISM OF SAMARIUM PHOSPHORESCENT SUBSTANCES BY MEANS OF STUDYING THE DECAY OF THEIR LUMINESCENCE

By W. LEWSCHIN and E. RICKMAN

(Communicated by S. I. Wawilow, Member of the Academy, 28.VI.1938)

As is well known, there are two classes of phosphorescent substances differing in the mechanism of luminescence of long duration (1). For substances of the first class the excitation comes to the displacement of the electron within the centre of luminescence, correspondingly to which the law of luminescence decay turns out to be exponential. For substances of the second class the excitation is accompanied by a complete separation of the electron from the centre of phosphorescence and the law of decay has a hyperbolic character. The majority of the most typical phosphorescent substances possessing long after-glow, e. g. zinc and alkaline-earth phosphorescent substances activated by small admixtures of heavy metals, belong to this class (2).

According to modern views on the process of phosphorescence the absorption of exciting light occurs for the substances of the second class in the lattice of the base substance, from which the excitation energy is further transmitted to the ion of the activator (3). The absorption spectra of these phosphorescent substances consist of a series of bands not connected with the absorption spectra of the activator itself, whereas the radiation spectra have the form of broad diffuse bands characteristic of the activator involved. A substantially different spectrum is given by the alkaline-earth phosphorescent substances activated by rare earths; their absorption spectra, similarly to those of all other phosphors of the second class, consist of a series of bands not connected with the activator, while their radiation spectra are composed of separate lines belonging to the tervalent ions of the activator. This peculiarity of the radiation spectra of phosphorescent substances activated by rare earths demanded for the establishment of the mechanism of their phosphorescence a special consideration of the laws of luminescence decay, since the discrete character of the radiation spectra is indicative of a weak interaction between the centres of phosphorescence and the surrounding medium. Besides, it is known that solutions of rare earths'salts are capable of luminescence possessing the properties of that of the substances of the first class, which together with the discrete character of the radiation spectrum suggested the possibility that the phosphorescing powders activated by rare earths also belong to this class.

The object of investigations was the decay of separate lines of CaSSm-phosphor of the following composition:

$$10\,CaO + 6\,S + 0.8\,NaF + 0.5\,(Na_2SO_4,\ Na_2HPO_4,\ Na_2B_4O_7) + 0.0003\,Sm$$

which gave two characteristic doublets of the tervalent ion of samarium: the yellow doublet (0.56 μ) and the red doublet (0.61 μ). Moreover, in the blue part of the spectrum of the investigated phosphor there was a rather broad weak band. The excitation spectrum of this phosphor consisted of four bands (maxima ~0.40, 0.35, 0.29 and 0.26 μ).

The separation of the red doublet and the blue band of radiation was achieved by means of appropriate filters, and that of the yellow doublet—by means of a double monochromator. The intensity of luminescence was measured with a special photometer by the «threshold stimulus» method (4).

For both doublets the measurements were carried out at three temperatures: 20, 100 and 200° C. The results for the yellow doublet are given in Fig. 1. During the measurements at temperature of 20° the brightness of luminescence of the phosphor in this doublet fell in the ratio of 1 : 46, at the temperature of 100°—in the ratio of 1 : 18 and at the temperature of 200°—in the ratio of 1 : 13. The results for the red doublet are given in Fig. 2. The

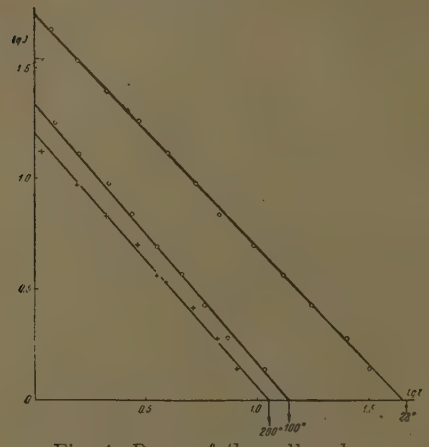

Fig. 1. Decay of the yellow doublet of CaSSm-phosphor.

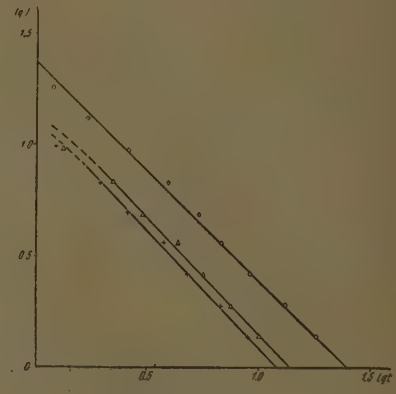

Fig. 2. Decay of the red doublet CaSSm-phosphor.

brightness of luminescence of the phosphor in the red doublet changed during the measurements at the temperature of 20° in the ratio of 1 : 18 and during the measurements at the temperatures of 100° and 200°—in the ratio of 1 : 9. For the blue-greenish band the measurements were carried out only at room temperature, and the brightness changed during the experiment in the ratio of 1 : 25.

As Figs 1 and 2 show, $\log J$ depends linearly on $\log t$, where J is the brightness of luminescence measured in arbitrary units and t is time in seconds. Thus, the law of decay proves to be strictly hyperbolic, and is expressed by the formula

$$J = At^{-\alpha} \qquad (1)$$

characteristic of zincic phosphors and other typical phosphorescent substances activated by heavy metals, the luminescence of which proceeds according to the bimolecular scheme. In the table are given the values of α characterizing the decay velocity of phosphorescence.

The Values of α for Different Temperatures

	20°	100°	200°
Yellow band	1.07	1.17	1.11
Red band	1.00	1.07	1.04
Light blue band	0.96	—	—

From this table we see that the rising of temperature does not practically have any influence on the decay constant. The observed tendency, common for both doublets, of the change of α with temperature, namely the acceleration of the decay when the temperature rises to 100° and the subsequent retardation when the temperature goes on rising to 200°, is so weakly expressed that it seems to be premature to subject it to a closer discussion.

From Figs 1 and 2, showing the straight lines $\log J = \log A - \alpha \log t$, we see that the rise of temperature, although not changing the course of decay (α remains constant), diminishes the value of A and, consequently, also the light sum L for the whole interval of decay, in which the approximative formula (1) is valid. The corresponding data for the change of A, or the radiated light sum L, with the temperature are given in Fig. 3. Curve I refers to the red doublet and Curve II to the yellow doublet. For $t = 20°$ C the value of L is taken to be 1 for both doublets. From Fig. 3 it follows that the first rise of temperature diminishes L rather abruptly, but a further rising has a much weaker effect. The doublets are not equally sensitive to the rise of temperature. The decline of brightness of the yellow doublet proceedes considerably quicker than that of the red doublet. Usually it is taken for granted that a rise of temperature accelerates the process of luminescence power exhaustion. This assumption which is certainly true for rather low temperatures is avowedly false in those cases when α has its limit value ($\alpha = 2$) already before heating. However, as we have seen, α may remain almost unchanged during the heating also in the case of comparatively small values ($\alpha \approx 1$). Since the decay of these phosphors proceeds at high temperature with a constant α, but with a smaller light sum, during the first instants of the heating of the phosphor excited at a low temperature, there must arise short flashing or quenching processes, during which the surplus energy absorbed by the phosphor at low temperature is emitted in form of radiation or is transformed into heat. From this point of view the flash phenomena do not amount to an acceleration of the fundamental process of luminescence power extinction, but to a short process of emittance of the surplus energy of the excited phosphor and to the transmittance of the course of its decay from the decay hyperbola corresponding to a large A to a hyperbola of the same order corresponding to a smaller A.

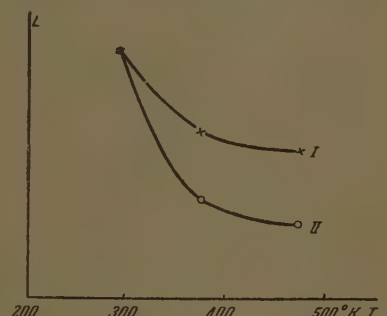

Fig. 3. Relative change in the light sum of separate doublets of Sm-phosphor as function of temperature.

In order to confirm the bimolecular course of the process, with the apparatus designed by V. Antonov-Romanovsky, an additional experiment has

been made under his participation, aiming at the determination of the initial brightness intensity of phosphorescence as a function of the intensity of excitation. It is known that for the bimolecular scheme of the process the initial brightness of phosphorescence is proportional to the square of the intensity of the exciting light ([5]). By investigating the decline of brightness of luminescence in the very first instants of decay and extrapolating of the obtained curve to $t=0$, it was found that at the change of the energy of the exciting light in the ratio of 1 : 4 the intensity of the original luminescence sank approximately in the ratio of 1 : 16, as it is demanded by the bimolecular scheme of the process.

The hyperbolic course of the phosphorescence decay of separate doublets of samarium phosphors established in the present paper and the quadratic dependence of the initial luminescence intensity from the intensity of the exciting light, as well as the earlier known independence of the absorption spectra of these phosphors from the activator, show that the alkaline-earth phosphors activated by rare earths must be considered, in spite of the discrete character of their radiation spectra, as belonging to the second class of phosphorescent substances, the excitation of which is accompanied by a complete tearing away of the electrons from the centres of phosphorescence.

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Received
1. VII. 1938.

REFERENCES

[1] S. I. Wawilow, Sow. Phys., 5, 369 (1934). [2] W. Lewschin and V. Antonow-Romanowsky, Journal Exper. and Theor. Phys., 4, 1022 (1934) and Sow. Phys., 5, 796 (1934). [3] R. Tomaschek, Phys. ZS., 878 (1932). [4] E. Brumberg a. S. Wawilow, C. R. Acad. Sci. URSS, III, 405 (1934). [5] V. Antonow-Romanowsky, C. R. Acad. Sci. URSS, II, 93 (1936).

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PHYSIKALISCHE CHEMIE

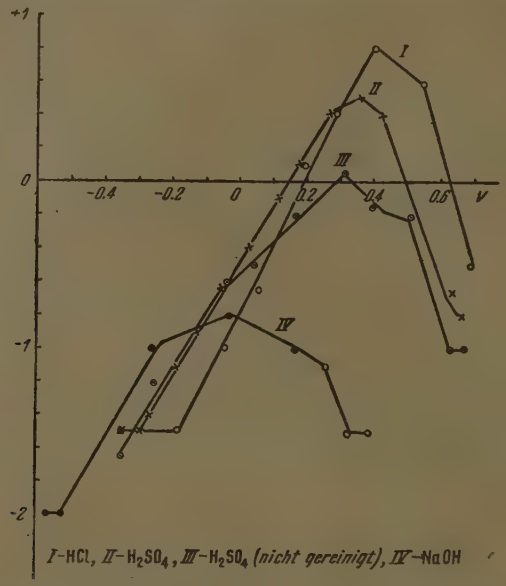
ELEKTROKINETISCHES VERHALTEN VON PLATIN BEI VERSCHIEDENEN POTENTIALEN

Von N. BALASCHOWA und A. FRUMKIN, Mitglied der Akademie

Billitzer (1) hat als erster versucht, das Zeichen der Ladung von Metallen aus ihrem elektrokinetischen Verhalten zu bestimmen. Seine Beobachtungen beziehen sich auf die Ablenkung von Metalldrähten im elektrischen Felde in Elektrolytlösungen sowie auf die Elektrophorese von kolloiden Teilchen. Er kam dabei zu dem Schlusse, dass für alle Metalle bei 0.47 V, gegen die Normalwasserstoffelektrode gemessen, ein universeller Nullpunkt der Ladung existiert. Das Potential wurde durch Zusatz von verschiedenen Oxydations- und Reduktionsmitteln variiert. Die meisten von Billitzer beschriebenen Versuche erwiesen sich aber als irreproduzierbar, ein Teil der mit Silber ausgeführten Messungen ausgenommen (2), hauptsächlich wohl deswegen, weil Billitzer die Schwierigkeit der Herstellung von reinen Metalloberflächen gar nicht beobachtete. Nachdem Freundlich und Smoluchowski zeigen konnten, dass das elektrokinetische und das totale Potential zwei voneinander unabhängige Grössen darstellen, begnügten sich spätere Forscher, insbesondere Coehn und Schafmeister (3), mit der Feststellung des Ausbleibens eines ausgesprochenen Zusammenhanges zwischen diesen Grössen. Sie nahmen an, dass die Edelmetalle sich wie typische negative Kolloide verhalten. Ein Versuch das ζ-Potential von Platin nach der Methode der Strömungspotentiale zu ermitteln ergab keine befriedigende Resultate (4).

In einer Reihe von Arbeiten aus unserem Laboratorium versuchten wir zu zeigen, dass trotz der Nichtidentität des elektrokinetischen und des totalen Potentials, zwischen diesen beiden Grössen ein weitgehender Parallelismus bestehen muss, wenn die Doppelschicht einen genügend einfachen Bau besitzt (Abwesenheit von kapillaraktiven und mehrwertigen Ionen, Abwesenheit von chemischen Veränderungen der Oberfläche). Unter diesen Bedingungen fällt der aus Adsorptionsmessungen bestimmte Nullpunkt der Ladung mit dem elektrokinetischen Nullpunkte zusammen. Diese Schlüsse wurden durch Versuche mit Kohlesuspensionen (5) und später auch durch Versuche mit Platinsolen (6) bestätigt. Mit diesen Objekten konnten aber direkte Messungen des totalen Potentials nicht ausgeführt werden. Wir entschlossen uns daher die Methode der Ablenkung von feinen Drähten auszunützen, die schon Billitzer und Coehn und Schafmeister gebraucht haben. Auch A. Dumanski und O. Dumanski (7) benutzten diese Methode für Messungen mit nichtmetallischen Oberflächen. In unseren Versuchen wurden besondere Massregeln getroffen um die Metalloberfläche rein zu erhalten. Nach der üblichen Reinigung wurde das Drähtchen aus einer reinen $PtCl_4$ Lösung mit einem sammtschwarzen Überzug von Platinschwarz bedeckt, so dass der Durch-

messer von 20 auf ca. 50 μ anstieg. Die untersuchten Lösungen (HCl, H_2SO_4 und NaOH, Anfangskonzentration $2.10^{-5}N$), wurden durch langdauernde kathodische Polarisation an einer grossen platinierten Elektrode gereinigt. Auch die Platinelektroden, die zum Anlegen des Feldes dienten, erlitten eine Reinigung durch anodische Polarisation. Auf Grund von anderen Arbeiten unseres Laboratoriums erwarteten wir, dass es nicht möglich sein wird eine Vergiftung einer glatten nicht platinierten Oberfläche in Gegenwart eines grossen Flüssigkeitvolumens vorzubeugen. Versuche mit nicht platinierten Drähtchen bestätigten diesen Schluss. Dagegen erwiesen sich die mit plati-



nierten Drähtchen erhaltenen Resultate als vollkommen reproduzierbar. Um den Einfluss von Giften unmittelbar zu untersuchen, wurden noch unter Beibehaltung der üblichen Bedingungen Versuche mit gewöhnlicher Schwefelsäure «zur Analyse» ausgeführt, die aber keiner speziellen Reinigung unterworfen wurde. Die Elektroden (Durchmesser 4.3 cm), die zum Auflegen des Feldes dienten, waren möglichst parallel zu einander in einem Abstande von 4 cm aufgestellt. Das Drähtchen, ca. 20 cm lang, befand sich genau in der Mitte; der Versuch zeigte, dass selbst eine kleine Asymmetrie die erhaltenen Resultate fälschte. Am Ende des Drähtchens ausserhalb des Feldes der Elektroden befand sich ein Glaskügelchen (Durchmesser ca. 1.5 mm). Auf die Elektroden wurde ein Potential von 160 V aufgelegt; unter diesen Bedingungen war die Potentialdifferenz zwischen verschiedenen Punkten in der Nähe der Metalloberfläche klein (Grössenordnung 0.1 V), was für die Anwendbarkeit der Methode sehr wesentlich ist. Die Ablenkungen wurden mittels eines horizontalen Mikroskops mit Okularmikrometer bestimmt. Die Ablesung erfolgte möglichst schnell und die Spannung wurde öfters kommutiert um den Einfluss der Elektrolyseprodukte auszuschalten. Um die elektroosmotische Bewegung der Flüssigkeit möglichst herabzusetzen, wurden die Elektroden von den Gefässwänden genügend weit entfernt, so dass die elektrische Feldstärke in der Nähe der Glasswand gleich Null war.

Die experimentelle Anordnung erlaubte das Potential des Drähtchens zu messen und die Lösung mit verschiedenen Gasen zu sättigen. Um einen

besseren Ausgleich des Potentials längs des Drahtes zu sichern, setzten wir in einigen Versuchen der Lösung eine gewisse Menge eines reinen Platinsols zu.

Die Versuche wurden wie folgt ausgeführt. Die Lösung wurde mit Wasserstoff gesättigt und als Ausgangspunkt das Wasserstoffpotential des Drähtchens und seine Ablenkung bei diesem Potential notiert. Dann wurde Stickstoff mit einem kleinen Sauerstoffgehalt durch die Lösung geleitet, so dass das Potential sich allmählich von dem Wasserstoffwerte zu dem Sauerstoffwerte bewegte, und die Ablenkungen des Drahtes bei verschiedenen Potentialen wurden gemessen. Typische auf diese Weise erhaltene Kurven sind auf der Figur dargestellt. Die Abszissen geben das Potential φ des Drähtchens gegen eine Normalwasserstoffelektrode gemessen, die Ordinaten, die Ablenkungen des Drähtchens—in Teilungen des Okularmikrometers an.

Kurve *I* gibt den Gang des ζ-Potentials in einer HCl-Lösung wieder. Wie ersichtlich ändert sich zunächst die Ablenkung bei einer Verschiebung des Potentials nach der Sauerstoffseite nicht. Dann wird das ζ-Potential weniger negativ und geht bei einem φ gleich 0.15—0.18 durch einen Nullwert. Bei positiveren φ-Werten ist das ζ-Potential positiv, erreicht ein Maximum bei ca. 0.4 V, fängt dann an zu fallen und wird schliesslich wieder negativ. In Schwefelsäure (Kurve *II*) beobachtet man einen ähnlichen Gang, nur sind die positiven ζ-Werte bedeutend kleiner. In nicht gereinigter Schwefelsäure (Kurve *III*) erfolgt keine Umladung. In alkalischer Lösung (Kurve *IV*) beobachtet man bei mittleren φ-Werten nur eine Verkleinerung der negativen Ablenkung. Der Absolutwert der negativen Ablenkung im Anfangsteile der Kurven liegt in alkalischer Lösung höher, als in saurer.

Die erhaltenen Resultate stehen in vollkommener Übereinstimmung mit den Messungen der Kataphorese von Platinsolen ([6]), sowie mit den Adsorptionsmessungen, die mit Platinelektroden bei höheren Konzentrationen ausgeführt wurden ([8]). Da bei den von uns verwendeten niedrigen Konzentrationen die spezifische Adsorption des Anions wohl vernachlässigt werden kann, können die von uns in Säuren erhaltenen Kurven am besten mit der Abhängigkeit der Adsorption vom Potential, die in einer angesäuerten norm. Na_2SO_4-Lösung beobachtet wird, d. h. in einer Lösung, die keine kapillaraktiven Anionen in hoher Konzentration enthält, verglichen werden. In dieser Lösung liegt der Nullpunkt der Ladung, d. h. der Übergang von negativer Ladung (Kationenadsorption) zur positiven (Anionenadsorption) bei ca. φ=0.15, in vollkommener Übereinstimmung mit unseren Resultaten. Bei weiterer Verschiebung nach der Anodenseite wächst zunächst die Anionenadsorption, fängt dann aber an in Folge der Ausbildung einer Oxydschicht zu fallen, so dass schliesslich die Elektrode wieder in ein Gebiet der bevorzugten Kationenadsorption übergeht. Qualitativ besteht auch in diesem Teile der Kurve vollständige Übereinstimmung mit den elektrokinetischen Resultaten. Ein quantitativer Vergleich wäre nur dann möglich, wenn die elektrokinetischen Versuche bei demselben pH ausgeführt würden, da die Lage des Oxydationsgebietes auf der Potentialachse sehr stark von dem pH der Lösung abhängt. Den flachen Anfangsteilen, die auf den in dieser Arbeit aufgenommenen Kurven sichtbar sind, entsprechen analoge horizontale Strecken, die im Wasserstoffgebiete auf Adsorptionskurven gefunden wurden. Ihre theoretische Bedeutung wurde schon früher diskutiert ([8]). Die in alkalischen Lösungen erhaltene Kurve *IV* zeigt, wiederum in vollkommener Übereinstimmung mit Adsorptionsdaten, dass in diesen Lösungen die Kationenadsorption an einer Wasserstoffelektrode in die Kationenadsorption an einer oxydierten Oberfläche stetig übergeht, ohne dass sich dazwischen ein Gebiet der Anionenadsorption, also positiver Ladung der Elektrode, einschaltet. Sehr interessant ist die mit nichtgereinigter Schwe-

felsäure erhaltene Kurve, welche zeigt, dass bei der Vergiftung der Oberfläche negative Gruppen (vielleicht As), welche von der Oberfläche fixiert werden, eine ähnliche Veränderung der Struktur der Oberfläche hervorrufen, wie die Ausbildung einer Oxydschicht.

Die beschriebenen Versuche, obwohl sie einen nur orientierenden Charakter haben, zeigen, dass die sehr einfache Methode der Drahtablenkung unter Beibehaltung bestimmter Bedingungen mit Erfolg zur Untersuchung des Zustandes von Metalloberflächen verwendet werden kann. Wir hoffen die Genauigkeit dieser Methode später wesentlich zu vergrössern. Elektrokinetische Methoden können nur dann auf Metalle angewandt werden, wenn die Grenzfläche Metall—Lösung genügend polarisierbar ist, die Stromlinien also parallel zur Metalloberfläche verlaufen. Man kann a priori behaupten, dass diese Bedingung in unserem Falle realisiert ist, wenn die H_2-Konzentration in der Lösung genügend herabgesetzt wird. Die von uns erhaltenen Kurven zeigen aber, dass sie auch in der Nähe des Wasserstoffpotentials realisiert wird, da sonst in diesem Gebiete ein starkes Abfallen des absoluten Wertes der Ablenkung auftreten würde.

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ZITIERTE LITERATUR

[1] Billitzer, Drudes Ann., **11**, 902 (1903); Trans. Am. Elektrochem. Soc., **57** (1930). [2] Palmaer, ZS. physik. Chem., **59**, 129 (1930); Freundlich u. Mäkelt, ZS. Elektrochem., **15**, 161 (1909); Garrison, J. Am. Chem. Soc., **45**, 37 (1923); Proskurnin u. Frumkin, ZS. physik. Chem., (A) **155**, 29 (1931). [3] Coehn u. Schafmeister, ZS. physik. Chem., **125**, 401 (1927). [4] Kruyt u. Oosterman, Proc. Acad. Wet. Amsterdam, **40** (1937). [5] N. Bach, Koll. ZS., **64**, 153 (1933); Kriworutschko u. N. Bach, ibid. 287. [6] N. Bach u. Balaschowa, Acta physicochimica, **3**, 79 (1935); Balaschowa u. N. Bach, ibid, **7**, 899 (1937). [7] A. Dumanski u. O. Dumanski, Koll. ZS., **66**, 24 (1934). [8] Šlygin, Frumkin u. Medwedowsky, Acta physicochimica, **4**, 911 (1936); Frumkin u. Šlygin, ibid., **5**, 819 (1936).

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PHYSIKALISCHE CHEMIE

METHODE ZUR BESTIMMUNG DES DIFFERENTIALPORENVOLUMENS PORÖSER KÖRPER

Von N. A. FIGUROWSKY

(Vorgelegt von Wl. A. Kistiakowsky, Mitglied der Akademie, 7.VI.1938)

Die Frage über die Dispersitätscharakteristik poröser Körper kann bis heute nicht als gelöst betrachtet werden. Die zur Porosität, bzw. Ultraporositätbestimmung, vorhandene Methoden ([1], [2]) geben kein klares Bild über die Grösse der Poren und deren Verteilung. Die Methode der Kapillarkondensation bei verschiedenem Druck ([3]), die auf der Gleichung von W. Thomson beruht, gestattet die Porengrösse zu berechnen und eine Differentialvolumenkurve zu ermitteln, erweist sich jedoch für grobporöse Körper, deren Porenweite einige Zehntel μ übertreffen, als unanwendbar. Zur Charakteristik dieser Art Körper ist man genötigt umständliche mikroskopische Messung durchzuführen, die aber kein vollkommenes Bild der Porenverteilung geben.

Die weiter angegebene Methode zur Gewinnung der Differentialporenvolumenkurven der porösen Körper ist auf die Verdrängung aus den Poren des betreffenden Körpers der einen Flüssigkeit (die im voraus eingeführt wird) durch die andere zurückzuführen. Beide Flüssigkeiten werden so gewählt, dass sie 1) das gleiche wahre spezifische Gewicht des porösen Körpers bei pyknometrischen Messungen aufweisen, 2) diesen Körper vollkommen benetzen und nahe Werte der Benetzungswärme ergeben, 3) eine unbeschränkte Gegenlöslichkeit besitzen, 4) eine Differenz der spezifischen Gewichte aufweisen. Als Beispiel Flüssigkeitpaare dieser Art dienen bei hydrophoben porösen Körpern (Kohle) Alkohol und Äther.

Füllt man die Poren eines körnigen Körpers (die von beiden Seiten geöffnet sind) vorher mit Äther und taucht man denselben sodann in Alkohol ein, so beobachtet man eine allmähliche Ätherverdrängung, deren Geschwindigkeit offenbar den Porenweiten proportional ist. Das Verdrängen wird in diesem Fall anscheinend nicht durch die Kapillarkräfte bestimmt, da der Meniskus auf der Grenze beider Flüssigkeiten fehlt (es ist auch ein umgekehrter Verdrängungsvorgang von Alkohol durch Äther möglich). Sowohl eine Diffusion, als auch ein Durcheinanderrühren, sind von geringer Bedeutung, da in Poren kleiner Weite (nicht über 100 μ) keine turbulente Strömung angenommen werden kann. Die Verdrängungskraft ist in diesem Fall offenbar nur durch die Differenz des spezifischen Gewichts beider Flüssigkeiten bedingt $(D-D')$.

An Hand der von Hagenbach (4) angegebenen Methode zur Erhaltung der klassischen Formeln von Poiseuille, findet man leicht die Abhängigkeit der Verdrängungsgeschwindigkeit von der Porengrösse für monodisperse Körper:

$$R = K\sqrt{\frac{L}{t}}, \quad \text{wo} \quad K = \sqrt{\frac{8\eta}{(D-D')g}}. \qquad (1)$$

Hier ist R—der Porenradius, L—dessen Länge, t—Dauer der vollständigen Verdrängung der Flüssigkeit aus dem monodispersen System, η—die Flüssigkeitsviskosität und g—Beschleunigung der Schwerkraft.

Gleichung (1) ist nur auf streng stabile Bewegung anwendbar. Es ist jedoch leicht nachzuweisen, dass das erwähnte Ergebnis auch bei beschleunigter Bewegung sich wenig ändert (für die Verdrängungsdauer von etwa 100 sek. verändert sich die Konstante um 0.0001 der Anfangsgrösse).

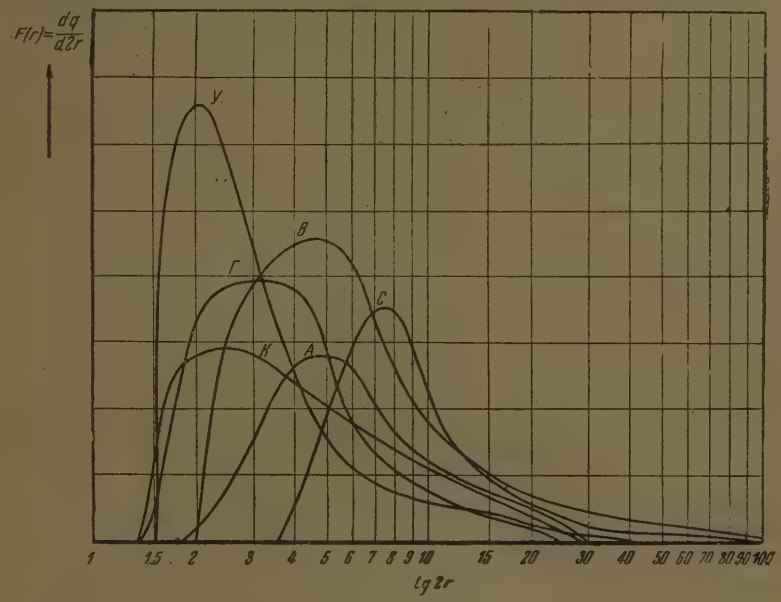
Die gewonnene Gleichung ist der Stokes'schen ähnlich (5), die zur Bestimmung der Teilchengrösse nach deren Fallgeschwindigkeit, bzw. deren Auftauchen (Suspensionen und Emulsionen), angewandt wird. Aus der Stokes'schen Gleichung ausgehend, hat Sven Oden (6) eine einfache Methode zur Gewinnung der Suspensionverteilung als Funktion ihrer Zersetzungsgeschwindigkeit angegeben. Es ist leicht anzusehen, dass die für monodisperse poröse Körper formelle Analogie ohne Schwierigkeiten auch auf polydisperse Systeme übertragen werden kann. Die Kurve der Verdrängungsgeschwindigkeit kann also ähnlich der Fällungskurve gedeutet werden (7,8) und nach der Verdrängungsgeschwindigkeit kann man folglich die relativen Porenvolumina polydisperser Körper verschiedener Grösse berechnen.

Zur Darstellung der Verdrängungskurven benutzen wir die bereits für andere Zwecke angewandte Glasmikrowage (9). Eine Probe körniger Kohle von bestimmter Korngrösse (ein Würfel, der aus dem angegebenen Muster ausgeschnitten wird, ist besonders gut geegnet) wird in ein kleines Glasgefäss mit Öffnungen am Boden gebracht. Die Kohle wird zunächst mit Äther unter Sieden gesättigt, worauf sie im selben Glasgefäss in ein grösseres Gefäss mit etwa 0.5 l Alkohol gebracht wird. Das kleine Gefäss wird sodann an einem feinen Platindraht am Balken der Mikrowage befestigt. Das gesamte System wird in ein Wasserthermostat gestellt. Gleichzeitig mit dem Eintauchen der mit Äther gesättigten Kohle wird die Stoppuhr in Gang gesetzt, wobei die Ausganglage der Wage möglichst rasch vermerkt wird. Die Lage des Balkens wird dann als Funktion der Zeit beobachtet. Die auf Grund der vorhandenen Zahlenangaben hergestellte Kurve wird berechnet, indem die bei den sedimentometrischen Berechnungen übliche Methode angewandt wird. Auf diese Weise wurde die Kurve der Porendifferentialvolumina ermittelt.

Wir haben verschiedene Proben aktiver Kohle untersucht. Die Abbildung stellt Kurven der Differentialporenvolumina für 6 Proben dar. Auf den Abszissen sind die äquivalenten Durchmesser (logarithmische Skala), auf den Ordinaten—die Verteilungsfunktion abgelegt. Aus der Abbildung geht hervor, dass sämtliche Kohlenarten je nach der Zusammensetzung recht verschieden sind, so z. B. zeigen Holzkohlensorten gegenüber den Nusschalenkohlen bedeutend weitere Poren auf. Aktive Birkenkohle (B) hat im Bereich der äquivalenten Durchmesser die grösste Porenweite (3—7 μ), während unaktive Birkenkohle (C) die grösste Porosität in den Grenzen 5—10 μ angibt. Kohlenarten, die durch Aktivierung von Pfirsichnussschalen (Y), Haselnusschalen (K), wie auch durch die Walnusschale (T) ihrer grobporösen Struktur nach ähnlich sind, zeigen im Bereich 1.5—5 μ die höchste Porosität. Die aktive Granuellen der Kohle zur Rückgewinnung verdünstender Lösungsmittel liefert im Bereich (3.5—7 μ) eine sehr wesent-

liche Porosität. Ein Vergleich der Kohlenverteilungskurven *B* und *C* gibt einen Begriff über die Veränderungen, die bei grobporösen Gefügen im Laufe des Aktivierungsprozesses vor sich gehen. Bei der Berechnung der gesamten Porosität ist das gesamte Porenvolumen der *B*-Kohle für 100% anzunehmen. Die Schnittfläche, die sowohl von der Kurve als auch von der Abszissenachse umgrenzt ist, drückt entsprechend das graphische gesamte Porenvolumen aus.

Zur Prüfung der Genauigkeit unserer Definition wurden mikroskopische Untersuchungen der Kohle *B* unternommen. Die Kohlenschnittfläche im Zelloidin wurde bei durchgehendem Licht verfolgt. Es ist festgestellt worden, dass die grössten Porenweiten der *B*-Kohle 150 μ betragen. Die Grösse der meisten Poren liegt jedoch zwischen 5—25 μ. Die Berechnung der Ver-



Kurven der Differentialporenvolumina verschiedener Kohlenproben.

drängungskurve für die *B*-Kohle ergibt, dass 70% des Porenvolumens auf Porenfraktionen mit 3.5—24.2 μ fallen. Somit ist ein gutes Übereinstimmen der auf beiden Versuchswegen ermittelten Ergebnisse festzustellen.

Die gegenwärtigen Kenntnisse über die Natur der aktiven Kohlen vermutet in diesen eine Existenz von sog. Ultraporen, deren Grösse sich nur sehr gering von denen der Moleküle unterscheidet. Die Ultraporen sind angeblich der gleichen Natur, wie die der groben Poren, d. h. der betreffende Körper wird von ihnen längs der gesamten Kerngrösse durchdrungen. Auf Grund unserer Ergebnisse sehen wir uns gezwungen andere Schlüsse über die Zusammensetzung der porösen Sorbenten zu ziehen, nämlich, dass die Ultraporen nichts anderes als Risse und Unebenheiten an der inneren Oberfläche der groben Poren darstellen. Unter diesem Gesichtspunkt wird die Aktivität der Sorbenten, wie auch der Katalysatoren, je nach der optimalen Grösse der groben Poren und je nach dem Rauheitsgrade, bzw. in Abhängigkeit von der Ausbildung der Risse, bestimmt.

Es wäre deshalb von Bedeutung den porösen Körper zweierlei zu bestimmen, mit anderen Worten sowohl einen Begriff einer scheinbaren als auch

wahren spezifischen Sorbentenoberfläche, wie auch eines Koeffizienten zur Darstellung der groben Porosität und des Rauheitsgrades zu geben.

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ZITIERTE LITERATUR

[1] М. М. Дубинин, Физико-химич. основы сорбционной техники (1935). [2] F. Krzeil, Untersuchung und Bewertung technischer Adsorptionsstoffe, 71—111 (1931). [3] P. Kubelka, Koll. ZS., **55**, 129 (1931). [4] Э. Гатчек, Вязкость жидкостей, 15 (1932). [5] Т. Сведберг, Коллоидная химия, 130 и далее (1930). [6] Sven Oden, Sedimentation Analysis and its Application; J. Alexander, Colloid Chemistry, **1**, 861—909 (1926). [7] H. Gessner, Die Schlämmanalyse, 31 u. folg. (1931). [8] К. А. Путилов, К теории и практике седиментометрич. анализа суспензий. Сб. под ред. П. А. Ребиндера, 60 (1933). [9] Н. А. Фигуровский, Зав. лаб., **6**, 335 (1937); Ж. пр. химии, **10**, 1142 (1937).

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PHYSICAL CHEMISTRY.

MOSAIC BLOCKS OF SILICATE GLASSES

By N. A. ŠIŠAKOV (N. A. SHISHAKOW)

(Communicated by Wl. A. Kistiakowsky, Member of the Academy, 22.III.1938)

The object of this communication is to give an interpretation of the phenomenon which may be easily observed in finest water suspensions of glass particles. In agreement with other modern data one is compelled to draw the conclusion that glasses are built of regular mosaic blocks and that a number of their properties can readily be explained on the basis of this idea.

The phenomenon may be briefly described as follows. If glass is ground into a comparatively fine powder and then shaken with water, its further spontaneous prompt disintegration (peptization) follows immediately. One becomes easily convinced of this fact by placing the suspension in a test-tube and observing the run of sedimentation. As by means of a simple grinding only comparatively coarse particles may be obtained, they will all gather at the bottom of the test-tube in a short time (not more than a few hours). However, not only after this lapse of time but even as late as in several days, the liquid above the sediment will still remain turbid, thus indicating that it contains much finer particles than those obtained by grinding. These fine particles are obviously due to the disintegrating action of water.

If such suspension is illuminated by a faint side light and looked against a comparatively dark background one can easily observe that the intensity of scattered light will alter in a discontinuous way with the depth of the layers, so that the whole liquid column appears to be divided into a number of alternating white and dark horizontal striations. This phenomenon seems to be similar to Liesegang's periodic reactions.

To give an explanation of this phenomenon we are forced to assume these finest particles to be not of any random size but only of some strictly definite sizes. If this assumption be true one should expect that some time after the beginning of sedimentation the upperest particles of different kinds will sink to a different depth, causing thus the appearance of visible borders between the layers of different concentrations.

True enough, these minute particles might be involved even in weak convectional currents, owing to which it seems scarcely probable to succeed in a stationary distribution of particles. Therefore all the experiments were performed in an isolated large dark room, remote from cold walls, windows, heating-radiators, etc.; any possible fluctuations of temperature were thus almost eliminated, and the interpretation of stratification, based upon the convection effect, was necessarily beforehand rejected. It proved necessary, too, to reject the original idea of the stratification as referable to the cathegory of Liesegang's reactions, since during the only possible reaction of formation of hydrates of silicium these reaction products might, obviously, be formed on the surface of grains as such thin layers that their influence upon the rate of precipitation of particles could be neglected.

The cause of this peptization of glass particles probably consists in that water is penetrating into the minute cracks present on the glass surface and within the latter and, thanks to the high hydrophility of glass and to the formation of absorbing layers, causes a weakening of the forces of attraction acting between the closely packed adjacent particles which then fall away from the remaining mother particles. That glasses are in fact heterogeneous in this sense has been of late recognized on the ground of their insufficient strength (1). The presence of cracks is also testified both by their ability to absorb large quantities of vapour and gases (2), and by direct microscopic examination of etching figures (3).

It seems now that some additional light can be thrown on the problem of constitution of glasses by the above phenomenon. The discontinuity in distribution of sizes of particles seems to show that there exists some regularity in the crack distribution in glasses, i. e. that glasses possess a true mosaic structure or, in other words, they consist of discrete minute blocks of similar size or are in simple ratio to each other. The surface etch-figures observed by Andrade (3) and which have in the majority of cases a quite regular character, seem to support this suggestion too.

To illustrate the phenomenon described a calculation of the «radii» r of the blocks of vitreous silica made by means of Stockes' equation is given below, assuming the viscosity $\eta=0.0102$ and the difference of the densities of particles and of water $\delta_1-\delta_2=1.22$. The observation of striations was started ten days after sedimentation set in.

h, cm	r, Å	Δr, Å	n	$\Delta r/n$
1.2	730			
2.7	1 095	365	5	73
3.0	1 154	59	1	59
3.4	1 229	75	1	75
4.1	1 349	120	2	60
4.5	1 413	64	1	64
4.8	1 460	53	1	53
5.2	1 519	59	1	59
5.6	1 576	57	1	57
6.0	1 632	56	1	56
6.3	1 672	40	1	40
6.9	1 750	82	1	82
7.3	1 800	50	1	50
8.0	1 884	84	1	84
8.8	1 976	92	1	92
		Average . . .		64 Å

In spite of a very rough computation it may be seen that the «radii» of particles corresponding to two adjacent bands always differ by the same value which in the given case is about 60—70 Å. The deviations from this rule in the case of differences Δr, given in the first and the fourth lines of the table, probably point to the intermediate layers having escaped from observation. I obtained in the same way from other experiments with vitreous silica the following mean values $\Delta r/n$: 79, 81, 76 and 80 Å; thus, the existence of this regularity seems to be beyond doubt. Such a regular increase of «radii» may result from the fact that the particles having a size exceeding the minimum are heterogeneous, but consist of these elementary mosaic blocks.

Similar experiments were done with some technical glasses, but although they gave quite similar data as to striation, no calculations of «radii» have been undertaken, since the densities of their particles were unknown. Out of other dielectrics examined in the same way, quartz, cristobalite and porcelain may be mentioned which also showed quite distinct striations. It may be of interest to note that while with quartz in the case of fresh suspension there was obtained a mean value $\Delta r/n$ about 60—70 Å, in the case of several month old suspension the average value from 22 striations proved to be about 20 Å. This points to the fact that the limit of peptization is attained not at once. Theire is, however, no doubt that it exists in reality, because no gradual transition to the upper transparent layer of the liquid was ever observed, the upper border being always sharply pronounced.

The values r calculated by means of Stockes' equation are quite approximate, owing to its validity in the case of spheric particles only. It seems, however, that the order of these values has been estimated correctly. In fact, according to Smekal's theory ([1]) «die Risslängen der tatsächlichen netzartigen Fehlstruktur dürften etwa mit 0.0001 mm und darunter anzusehen sein», whilst in my eight experiments with vitreous silica the values from 710 to 2530 Å for the elementary mosaic blocks were obtained. It has been also shown by Krishnan ([4]) by means of Raman spectra investigation that the molecular complexes in optical glasses are not smaller that the wave length, i. e. they have also the dimensions of the order of a few thousands of Ångström units.

In order to estimate the dimensions of pores, i. e. the thickness of empty spaces between adjacent mosaic blocks, we may take the case of vitreous silica where the difference between the structure density of the cristobalite crystals (2.31) which it probably consists of ([5]) and its apparent density (2.20) may be used. Assuming that the mosaic blocks are cubic in shape, we obtain in this case the following expression for the thicknessi d and the linear dimensions of mosaic blocks: $d=0.017\ \lambda$.

As the smallest blocks of vitreous silica observed in my experiments have $\lambda=2\ r=710$ Å, the mean thickness of the cracks will thus be 12 Å. Such a value enables us to explain many experimental facts, for instance: the permeability of glasses for molecules of small dimensions (water, nitrogen, oxygen, alcohol, ether), the penetration of ions due to electrolysis of glasses and, inversely, the non-permeability of glasses for long aliphatic chains (oil) and highly associated molecules of glycerine under high pressures ([6]). It seems, however, that the above thickness of pores is not constant and that it may vary in dependence on some or other factors, for it would be, otherwise, difficult to explain some facts, as, for instance, the increase of mechanical strength of glasses when testing them in a vacuum or heating ([7]), the decrease of tensile strength due to ageing ([8]) and other facts.

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REFERENCES

[1] A. A. Griffith, Phil. Trans., A, **221**, 163 (1920); A. Joffe, The Physic of Crystals (1929); A. Smekal, Glastechn. Ber., **13**, 141 (1935). [2] R. Bunsen, Wied. Ann., **20**, 545 (1883); H. A. Шишаков, Ж. пр. физ., **4**, 37 (1927) [3] E. N. da C. Andrade a. L. C. Tsien, Proc. Roy. Soc., **159**, 346 (1937). [4] R. S. Krishnan, Proc. Inst. Acad. Sci., A, **3**, 211 (1936). [5] N. A. Shishakow, Nature, **136**, 514 (1935); H. A. Шишаков, Ж. Т. Ф., **5**, 1834 (1935). [6] T. C. Poulter a. R. Wilson, Phys. Rev., **40**, 877 (1932). [7] P. A. Rehbinder, ZS. f. Phys., **72**, 191 (1931); Bull. Acad. Sci. URSS, sér. chim., № 5, 639 (1936) S. Schurkow, Phys. ZS. Sow. Union, **1**, 123 (1932); A. Smekal, loc cit.; O. Reinkober, Phys. ZS., **32**, 243 (1931); **33**, 32 (1932); **38**, 112 (1937). [8] A. A. Griffith, loc. cit.

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CHEMISTRY

ON CHARACTERISTICS OF THE PHENOMENA WHICH TAKE PLACE AT THE BOUNDARY LINE BETWEEN TWO SOLID PHASES IN AN INDIFFERENT LIQUID MEDIUM

By M. A. ILJINSKY, Honorary Member of the Academy

Freundlich in his introduction to the first edition of the book «Die Kapillarchemie» remarked that «A study of the system «Solid on solid» can be of no considerable interest, since the little mobility of the particles of solids hinders the display of the characteristic phenomena which take place at the boundary line between these two phases».

These considerations are correct only inasmuch as we have to deal with two solid phases in a gaseous surrounding. The introduction of an additional indifferent liquid medium which dissolves practically neither of the solids, makes it possible, as it has been shown for the first time by the author of this paper (1), to set quantitative experiments with the colloidally compact and powder-like bodies suspended in an indifferent liquid medium. Phenomena of this kind had been frequently observed before (2, 3), without being given proper interpretation.

It is known that coarse suspensions are converted into a state of fine dispersion upon introduction of colloidal solutions of glue, gum, etc. Analogous phenomena were observed when treating salts of heavy metals with alkalies in the presence of prosalbinic or lisalbinic acids. The colloidal particles of the alkaline salts of these acids envelop the hydroxides produced converting them into colloidal state. The purple of Cassius is an adsorption compound of colloidal gold and colloidal stannic acid in an aqueous medium. Comparing colloidal solutions to finely divided suspensions of solid particles, we have in this case to deal with two solid phases in an indifferent liquid medium too.

A study of such three-phase systems—liquid as a dispersing medium and two solid dispersed phases—offers considerable experimental difficulties.

A general case—two dispersed solid phases A and B and one liquid dispersion medium C—is too complicated. To begin with we have chosen fibre for one phase—an example of a compact solid easily wetted with water throughout its thickness—leaving one variable only for such a solid dispersed in a liquid phase as oxides of heavy metals, sulphur, carbon, inorganic pigments, organic dyestuffs and their semi-manufactured products.

Such a restriction of the field of the phenomena considered permits of approaching the display of molecular forces not only from the qualitative but also from the quantitative point of view. The dyeing of the fibre with an adsorbed dyestuff considerably facilitates the observation of the

reactions which take place, namely the degree of completeness and the character of deposition of the dyestuff on fibre, the influence of pH of the medium, the temperature, etc.

The following statements have been gradually developed ([4],[5],[6],[7]):

1. A perfect possibility to register qualitatively and quantitatively the occurring phenomena.

2. A selective character of the affinity of suspended solid particles with the fibre.

3. While the crystalline or coarsely dispersed materials (indigo, β-naphthol, hydroxides of iron or lead) suspended in a slightly acid aqueous medium are completely or almost completely deposited on fibre within 5 minutes, the deposit being d u r a b l e t o w a t e r, other materials, chiefly those having a colloidal nature (sulphide, dyes, complexes of azodyes, insoluble in water), pass from aqueous suspensions to the fibre rather incompletely, forming with the latter not durable or insufficiently durable to water complexes of purely mechanical character.

T h e p r e s e n c e o f c o l l o i d s (glue, gelatine, gumarabic, starch) p r e v e n t s t h e d e p o s i t i o n o f t h e m a t e r i a l from an aqueous medium onto the fibre. The latter is completely or almost completely decoloured by means of a simple washing with water.

4. Consequently the durable to water complexes, f i b r e—s o l i d, are not durable to the colloids; the fibre is completely or almost completely decoloured when treated with soap solutions.

5. A considerable influence on the deposition of solid particles onto the fibre has pH of the aqueous medium, irrespective of a larger or smaller affinity of the solid to the fibre.

The maximum display of affinity of the solid to the fibre takes place in a slightly acid medium. The depositions of the solid onto the fibre goes on less vigorously from neutral aqueous suspensions; it is even more difficult in an alkaline medium.

6. The amount of the aqueous dispersion medium is of considerable importance for the extractibility of the solids by fibre. Thus, the length of the bath should not exceed in the case of silk the ratio 500 : 1 to the weight of fibre, for wool—50 : 1, cotton—5 : 1 to permit of a complete removal, say, of indigo from an aqueous suspension of the same. The dependence of the degree of extractibility of the solid particles by fibre on the length of the bath has a paradoxical character, since the complexes fibre-solid once formed are practically durable to water (irreversible adsorption).

7. The durable to water dyestuff-fibre complexes, formed by depositing the dyestuff on fibre, are in the absence of colloids no ready dyeings, but present merely a primary labile superficial not durable to soap phase of the dyeing.

In order to obtain the final stable phase of dyeing throughout the fibre, the labile complexes have to be subjected to certain additional procedures. Thus, in the case of substantive or acid dyestuffs this will be an additional steaming. The dyeing with indigoids requires an additional treatment with alkaline hydrosulphite solution. The labile complexes of the para-red type (insoluble azodyes) fail to be converted to the stable phase.

8. The facts above throw a new definite light on the theory of dyeing too. Every process of dyeing goes on through a labile phase, never observed by anybody before us, while the processes of dyeing are carried out usually with an aim to obtain immediately the second stable phase without waiting for the final formation of labile complexes.

Yet, as shown by direct observations, the procedure of dyeing in two phases presents essential technical advantages: considerable economy of time,

and consequently, of the apparatus; a considerable economy of steam leading besides to a better preservation of the fibre, particularly of that of animal origin; the dyeing of the fibre throughout its thickness, etc. These paradoxical phenomena—better dyeing results obtained by means of suspension than with solution—are to be accounted for by the fact that colloidal water-fast films of a stable phase are formed on the surface of the fibre when the latter is dyed with an aim to obtain directly the second stable phase, with a premature raising of the bath temperature. These films prevent the farther penetration of the dyestuff deeper into the fibre.

Wool fibre. The phenomena of adsorption of insoluble in water acid-adjective dyestuffs have been subjected to detailed study at the Institute for Scientific Research in Wool Industry ([11]). The following relations have been found:

1. The degree of dispersion of the particles of dyestuff suspended in water exerts no material influence on the nature of the deposition of the dyestuff on the fibre.

In this case the finely or coarsely dispersed suspensions of the dyestuff behave in the same way. The particles of the dyestuff are deposited at first on the surface of the fibre as separate dots very close to each other, and giving the misleading impression of a complete evenness of colouring.

2. When the specimen is kept under moist conditions, still more rapidly when warmed, the procedure of dyeing assumes another character: the particles of the dyestuff react with the fibre. Individual particles of the dyestuff cease to be observable either on the surface or in the thickness of the fibre—intramolecular forces disperse the coarse particles deposited on the surface to the state of a finest division; the dyestuff penetrates simultaneously into the fibre, resulting in a complete and even dyeing throughout the fibre.

No less, if not greater interest has the study of the phenomena of adsorption of a dyestuff by wool from solution. The procedure of dyeing of wool with a solution envolves an intermediate formation of labile adsorption phase, similar to the case of insoluble dyestuffs. Depending on the nature of the dyestuff, some temperature increase may be necessary for a more or less complete deposition of the dyestuff onto the fibre. The transition of the labile surface-dyeing phase to a stable one, involving a complete throughout dyeing of the fibre, requires a more or less prolonged boiling.

On account of both great economic importance and theoretical interest of the accelerated adsorption dyeing of animal fibre with soluble acid and acid-adjective dyestuffs, the experiments undertaken by us should certainly be prolonged first of all in the following two directions:

1. Elucidation of the function of constitution and physical nature of the series of soluble acid and acid-adjective dyestuffs, as to their aptness to the two-phase adsorption dyeing of animal fibre.

2. Synthesis of new representatives of the series of acid and acid-adjective dyestuffs equal to the requirements set before the new method.

Cotton. Less sharply expressed display of chemical molecular forces in the vegetable fibre, compared to the animal one, requires the use of very finely divided suspensions of the dispersed dyestuff, to succeed in an even deposition.

The studies undertaken by R. S. Osipova and T. N. Shkurina at the Cotton Institute for Scientific Research have outgrown the stage of laboratory experiments and proceed with success on an industrial scale ([12]).

Neither wool nor cotton research can be regarded as completed. It is necessary to carry on the theoretical investigations undertaken at the 1st Moscow University ([7]) and the Karpoff Institute ([8]) to make clear the nature of the adsorption processes involved—the influence of the medium, of the presence

of electrolites, of the inhibiting action of organic colloids on the degree of deposition of the solid dispersed in water onto the fibre, etc.

Practical work should be done to extend the application of the adsorption method to other kinds of fibre—natural silk, viscose, acetate silk, compound fibre, etc.

Aside from its purely scientific interest, the study of molecular forces displayed at a boundary line between two solids may throw light on other problems beside that of dyeing.

Among others can be cited:

1. Ascertaining of the nature of solid solutions of metals and oxides in silicates.

2. Formation of geological stratifications and seams.

3. The nature of assimilating by the soil of solid fertilizers insoluble in water.

It should be mentioned as well that the treatment of phenomena that take place at the boundary between two solid phases would extend the domain of capillary chemistry by 25%.

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REFERENCES

[1] M. Iljinsky, ZS. f. angew. Chemie, 354 (1914). [2] Любавин, Технич. химия, 5, ч. 1, 640. [3] С. Виньон, II, 156 (1909). [4] Герм. патенты: № 270,520, 275121, 275570, 28704, 289314, 275103, 283716, 289908, 289314. Англ. патенты: № 27391/1912, 28656/1912. Амер. патент № 735930. Рус. патент № 55553. [5] Сов. охранит. свидетельства № 73921/1919, 73953/1920, 73921/1920. [6] М. А. Ильинский, Технико-эконом. вестник, № 3 (1924). [7] М. А. Ильинский, М. В. Гавардовская, М. Б. Турова, Поляк (1926). [8] М. А. Ильинский и В. В. Козлов, Ж. Р. Ф. Х. О. **62**. [9] М. А. Ильинский и Р. Бурштейн, Тр. Ин-та Карпова (1928). [10] М. А. Ильинский, Хлопч.-бум. промышленность, № 11/12, 46 (1935). [11] С. С. Рахлина, Шерст. дело, январь (1937). [12] Р. С. Осипова и Т. Н. Шкурина, Хлопч.-бум. промышленность, № 4 (1937).

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CHEMISTRY

STRONTIUM AS A CHEMICAL INDICATOR

By V. I. NIKOLAIEV and S. K. KALININ

(Communicated by N. S. Kurnakov, Member of the Academy, 20.VI.1938)

The quantitative determination of those elements which are present in very small quantities in natural waters is at present becoming extremely necessary in solving geological and geochemical problems, as well as for the physico-chemical exploration of salt deposits.

It has already been established (1) that such elements as potassium, bromine and boron may be used as indicators.

In this work we shall show that strontium can also be used as an indicator.

In appraising the analytical data on strontium, when analysing salt waters, it is necessary first of all to take into account the following considerations and data.

We are able to conjecture the composition of the concentrated brines of the Permian Sea, at a time when the latter was drying out, from the composition of those brines which are produced when the Permian salt deposits are redissolved in water, as is taking place now under natural conditions, for instance in Inder Lake.

The following peculiarities constitute the difference between the Permian Sea and the Caspian Sea of to-day.

Permian brines have very little magnesium sulphate and their coefficient of metamorphization is 0.1. Caspian waters and strong brines are, on the contrary, quite rich in magnesium sulphate. Their coefficient of metamorphization may be as high as 5. The KCl content of the brines which are produced when the Permian salt deposits are dissolved is considerably greater than that of Caspian waters. Similarly, a considerable increase in the content of bromine and boric acid in comparison with the brines of the Caspian Sea was noticed in the Permian brines (2).

After having made many spectroscopic determinations of strontium in the waters of the Caspian Sea* with a quartz spectrograph (manufactured by The Adam Hilger Co) we make bold to estimate the content of strontium ion in the residue left from evaporation equal to 0.01%. The residue left after brines of 26° Bé in strength from Inder Lake had been evaporated were found to contain 0.07—0.08% $Sr^{\cdot\cdot}$, i. e. 7—8 times more than in the residue from Caspian waters.

According to the data given by Wolfman (3) who studied the solubility of $SrSO_4$ in water at various temperatures, this salt belongs to the class of slightly soluble substance, viz.

The solubility of $SrSO_4$ in 100 g of water:

at 5°	10—12°	20°	30°	50°
0.0983	0.0994	0.1479	0.1600	0.1629

If NaCl is present in the water, the solubility of $SrSO_4$ increases up to 0.219%.

* The water test was taken several kilometres from the Karabugas Bay, in May 1937.

Similarly, under the influence of considerable quantities of KCl (18%) the solubility of $SrSO_4$ increases to 0.251% ([4]).

On the other hand, under the influence of magnesium sulphate and calcium sulphate the solubility of $SrSO_4$ should decrease.

A comparison of these data on the solubility of $SrSO_4$ leads us to expect low concentrations of strontium in waters of Caspian origin, which are rich in magnesium and calcium sulphates, and high concentrations in waters and brines of Permian origin, which are rich in KCl+NaCl.

Thus, we now have the right ro regard the strontium ion as a new indicator which, together with the other indicator elements. $K^{\cdot}$, Br and also boron, makes it possible for us to conjecture the nature and genesis of the salt deposits and natural waters and to distinguish Caspian salt deposits from those of Permian origin.

We may regard concentrations of $Sr^{\cdot\cdot}$ ions of over 0.02% as being characteristic of salt deposits of distinctly Permian origin; while concentrations not over 0.01 to 0.02% may be taken as characteristic of Caspian deposits.

Our analysis of a number of natural waters for $Sr^{\cdot\cdot}$ ions are given in the table.

Analyses of various waters and brines for $Sr^{\cdot\cdot}$ ions

Origin of the brine	$Sr^{\cdot\cdot}$ in weight % of the residue after evaporating
Water from the Caspian Sea (a few kilometers from Karabugaz Bay)	~ 0.01
Brine (26° Bé) from Inder Lake	0.07—0.08
Balkhash Lake	0.001—0.003
Spring Gazovyi at Inder Lake	0.02—0.04
New Ashche-Bulak Spring at Inder Lake	0.02—0.04
Spring near Kara-Chungul Lake, feebly ejecting chalybeate waters	0.07—0.08
Brine from the lake between the Iskine mines and hamlet Rakushy	0.001—0.002
Water from salt-well № 167 in the Kosh-Chagyl mine	0.006
Spring № 1 at the Alma-Arasan resort (Zaïlisky Alatau)	0.001—0.002
Spring № 2 at the Alma-Arasan resort	0.001—0.002
Iskine salt lake	0.001—0.002
Uruskul Lake, Jarkent region	0.001—0.002

These analyses show that the percentage concentration of $Sr^{\cdot\cdot}$ ion in reservoirs so obviously of Caspian origin as Lake Balkhash, is in the thousandth parts of one per cent. On the other hand, the $Sr^{\cdot\cdot}$ content in the brine of Inder Lake and in the water of springs around Inder Lake (Gazovyi Spring, New Ashche-Bulak Spring, and also the spring near Lake Kara-Chungul) is from 0.02 to 0.08% which shows their connection with Permian salt deposits.

The spring near Lake Kara-Chungul is especially noteworthy and so, consequently, is the whole district around this lake. We found that the dry residue left after evaporating water from this lake contained 0.07—0.08% $Sr^{\cdot\cdot}$ which is exactly the same as the strontium content in Inder Lake.

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REFERENCES

[1] V. I. Nikolaiev, C. R. Acad. Sci. URSS, XV, № 8 (1937). [2] V. I. Nikolaiev, O. K. Janatieva a. V. D. Poliakov, Bull. Acad. Sci. URSS, sér. chim., № 4 (1937). [3] J. Wolfmann, Oester, Zeit. Zuckerind., 25, 986 (1897). [4] A. Virck, Zeit. anal. Chem., I, 473 (1862).

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CHEMISTRY

DEPOSITING OF METALS ON MONOCRYSTALS

By K. M. GORBUNOVA

(Communicated by Wl. A. Kistiakowsky, Member of the Academy, 16.VI.1938)

According to the data met with in literature, various authors have observed that in some cases the metal being deposited corresponds in structure to that of which the cathode is made. And although a number of authors (1) have been using monocrystallic cathodes, the conditions under which such a correspondance appears and the degree of perfection of the reproduced structure still remain a problem unsolved.

While studying the structural interaction between the basic cathode and the metal being deposited electrolytically, we succeeded in making a number of quite interesting observations regarding the depositing of a metal on a monocrystallic zinc surface.

For our investigation we designed an apparatus which is represented schematically in Fig. 1. The electrolyte comes from cylinder *B*, washes the anode* *A* and passes through a tube *T* to an opening *O*, 3 mm in diameter. A drop of liquid, coming from the opening, came into contact with one of the faces of the zinc monocrystal cathode *K*. In this way it was possible to get a single crystal face into the electrolytic unit without using insulators.

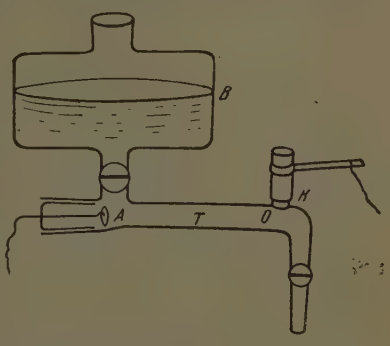

Fig. 1.

Faucets were used to control the velocity of flow of the electrolyte, which was widely varied in various experiments**. An aqueous solution of very pure zinc sulphate (2) was electrolyzed, as a result of which quite perfect hexagonal tablets were deposited on the face (0001). These tablets were exactly orientated in relation to one another and in relation to the initial crystal face. Fig. 2 shows the surface of such a deposit. At first the electrolytically deposited zinc reproduced the surface of the face as a whole, but then, as the electrolysis went further, the hexagonal tablets shown in the

* The anode was also a zinc monocrystal.

** In a number of experiments the electrolyte flowed through an opening toward which the cathode was moved.

microphotograph appeared; however, the latter were so perfectly orientated that as a whole they may be considered a mosaic monocrystal union.

When we used an artificial (polished) and etched surface of a monocrystal (perpendicular to the base) as cathode, the deposit turned out to consist of cry-

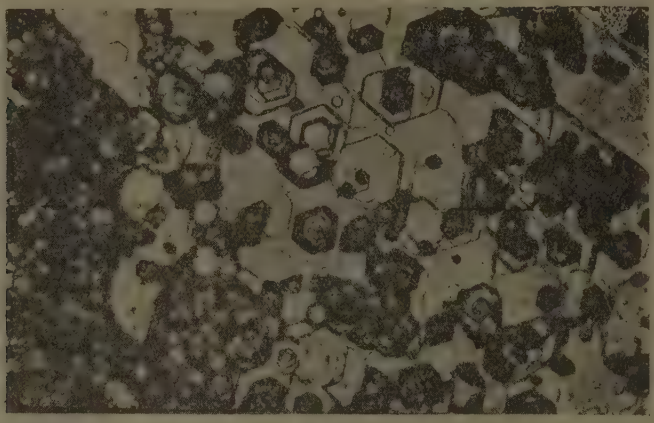

Fig. 2.

stallite unions, which mirror the orientation of the basic crystal. The thickness of separate layers (ledges) of the crystallite unions is somewhat less than that given by Straumanis (2) for crystals obtained through condensing vapour and equals approximately 0.4 μ. The edges of the crystals on the

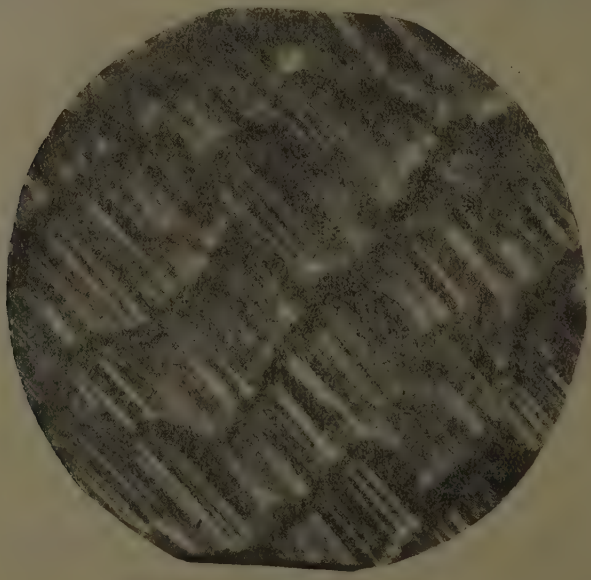

Fig. 3.

surface are inclined at an angle of approximately 22° to the plane of the cathode, which corresponds to the incline of an artificial surface of the cathode to the face of the monocrystal prism.

These experiments on the electrocrystallization of zinc show that if the solution is absolutely free of impurities and the metal surface absolutely clean, the structure of the lining is most likely to be reproduced during electrocrystallization. These experiments are very easily reproduced under the conditions described*. The length of time a freshly-cut surface of the crystal was kept in air had no effect on the results of the experiment.

The possibility of the structures corresponding when a different metal is deposited, and the influence of the specific conditions of electrocrystallization on the process are of special interest. For this reason we conducted some experiments on depositing metal on the face of a zinc monocrystal by dipping the latter into a solution of a salt of a nobler metal. While depositing cadmium (from its sulphate and halides) we could clearly observe, by using a microscope, the layer-like growth of the faces, that is typical

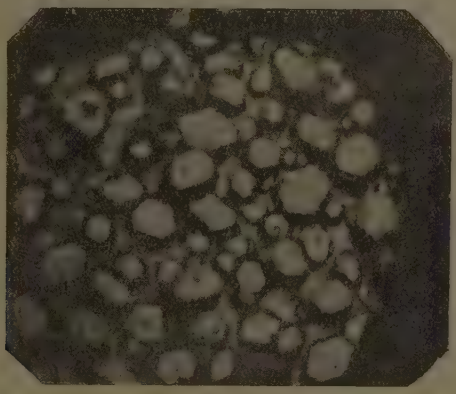

Fig. 4.

of the phenomena of the electrocrystallization of metals: a «bi-dimensional germ», originating at the edge of the face spread all over it which could be seen from the motion of the line bounding this layer. Thus became manifest that the mechanism of the growth of metallic crystals in electrocrystallization is similar to that which takes place when a metal is displaced from solution by a less noble metal. We were unable to establish a simple relation between the orientation of the cadmium crystals deposited and the basic zinc face at the concentrations studied.

Besides the experiments with cadmium we conducted some experiments in displacing and observing the crystallization of metallic copper, tin, lead and thallium onto the face (0001) of a zinc monocrystal. The thallium deposits were of special interest; they consisted of hexagonal (more or less perfectly developed) flat prisms, strictly orientated in relation to the zinc (0001). The primary thallium layer is a continuous compact formation. Fig. 4 shows a layer of crystals which has risen as a result of quick growth on top of the primary continuous layer. There are prisms of the first and second kinds. Correspondingly orientated branchy dendrites formed on the side faces of the monocrystallic zinc disc. The continuity of the coating and the strict orientation of the crystals on the face (0001) make it quite evident that this face does not diffuse.

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* Neither change in current density (from 35^{-3} to 10^{-30} A/cm^0) nor changes in the thickness of the deposit (from 30 to 110 μ) and the rate of flow of the solution affected the similarity between the structure of the deposit and that of the cathode.

From these experiments it is clearly seen that when one metal is deposited on the surface of another, crystallizing in the same system, the structures of these metals correspond, although the lattice parameters differ sufficiently (thus, for Zn $a=2.66$; $b=4.966$; for Tl $a=3.45$, $c=5.52$).

A more detailed report of the results of this investigation, including the problem of the polarization of the cathode in case of a crystal growing in a definite direction, is being prepared for print.

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REFERENCES

[1] Tamman u. Straumanis, ZS. anorg. Chem. **175**, 131 (1928); Setzura Takeyama, Mem. Coll. Science Kyoto Imp. Univ. (A), **13**, 363 (1930); Thompson, Proc. Roy. Soc. (A); **133**, 1 (1931); Kreuchen, ZS. phys. Chemie, **155**, 161 (1931); J. Chiout, J. Inst. Metals, **39**, 595 (1938); Erdey-Gruz, ibid. (A), **172**, 157 (1935); Erdey-Gruz u. Kardos, ibid. (A), **178**, 255 (1937); Erde-Gruz u. Frankl, ibiq. (A), **178**, 266 (137); Mikhalev, Bull. Acad. Sci. URSS, série chim., 789 (1937). [2] Straumanis, ZS. phys. Chemie (B), **13**, 316 (1931).

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GEOCHEMISTRY

ON THE PROBLEM OF THE ORIGIN OF KAOLIN

By P. A. ZEMIATCHENSKY, Corresponding Member of the Academy

While the origin of kaolins from potassium-feldspar rocks may be regarded as firmly established, only the details of the process being insufficiently clear, the formation of kaolins from calc-sodium-feldspar rocks is far from being explicit and there are as yet known but few deposits of such kaolins.

Still less studied is the part played by micas in the formation of kaolin in the course of weathering. In fact, it may be said that in literature (geological, mineralogical, ceramic, etc.) this problem has been given little attention (Schimmer, K. D. Glinka). However, more detailed and thorough investigations of the kaolin deposits of the Union indubitably point to certain large kaolin deposits as having originated from granite-gneisses with a complete transformation not only of feldspar but also of mica to kaolinite. Such is the Glukhovetz deposit at the Ukraine well-known by its vast reserves.

A study of the alterations and transformations taking place in such widespread minerals as micas is surely of considerable interest from the viewpoint of mineral and rock weathering processes in general.

Studying the kaolin deposits of the Ukraine the present author has had the opportunity of observing certain transitional stages of granite-gneisses weathering to kaolinization. This circumstance has been mentioned by the author in his description of the kaolin deposit in the neighbourhood of the village Blagodatnoye (1). Here, in a cave serving for primitive kaolin mining the gradual transitions of granite-gneiss slightly altered by weathering to kaolin may be seen in the fresh wall of the cut. Accumulations of a greenish scaly mineral clearly visible upon the white background of weathered feldspar, have been observed in the transitional zone, where the shape of the grains and the cleavage of feldspar have still been retained in spite of a complete transformation of the feldspar to kaolinite.

The above accumulations are of irregular shape, elongated in the direction of the lamination of the granite-gneiss, and consist of a jumble of rather large angular quartz grains, white grains of weathered feldspar, occasionally mixed with irregular grains of yellow-coloured opal while the greatest part of the segregations is formed by a greenish mineral having the shape of scales of different size. The largest lamellae are bent.

The chemical composition of the material carefully sorted out under a magnifying glass was as follows:

SiO_2	55.42	0.906		10.06
TiO_2	Traces	—		—
Al_2O_3	21.02	0.206	0.227	
Fe_2O_3	3.42	0.021		2.5
FeO	1.61	0.022		
CaO	None	—	0.092	1
MgO	2.82	0.07		
K_2O	7.04	0.090		1
Na_2O	0.40	—		—
H_2O on ignition	5.77	0.320		3.6
H_2O hygroscopic	2.86	—		—
	100.36	—		—

The above chemical composition may be represented by the general formula

$$K_2O\,(Mg,\,Fe)O \cdot 2.5\,(Al,\,Fe)_2O_3 \cdot 10SiO_2 \cdot 3.6H_2O \text{ or}$$
$$2\,K_2O \cdot 2RO \cdot 5R_2O_3 \cdot 20SiO_2 \cdot 7H_2O.$$

This formula shows the general composition of the greenish mineral; it cannot pretend to express the chemical characteristics of the given mineral, since the green lamellae present a mixture of different weathering products as will be shown further.

The specific gravity determined by means of a pycnometer at 15° C was 2.503. Hardness similar to that of talc. On ignition rapidly acquires a brownish colour due to the disintegration and evolvement of free ferric oxide; at the same time a partial cacking takes place. The colouring of different particles is not uniform, which points to a different Fe-content of these particles. Optically they are characterized by the following features.

In ordynary light the lamellae seem to be yellow-greenish; in polarized light they display a slight pleochroism (or, possibly, absorption). An abundance of small isometric inclusions is noticeable (occasionally rectangular outlines are marked).

Optically uniaxial (probably, biaxial with a very small axial angle), negative. The refraction index in the majority of scales is 1.557 ± 0.003. Lamellae with a somewhat higher refraction index (1.568) are of rare occurrence.

A detailed study of the mineral under microscope shows that in ordinary passing light the particles seem to be homogeneous while in polarized light between crossed nicols they fall apart to individual indistinct fields: dark and of different shades of grey of the 1st order; in the latter a fibrous structure, sometimes a fan-shaped one is observed. In the majority of lamellae the crystalline lattice seems to be considerably distorted which results in the disappearance of clear and regular optical phenomena. In some lamellae only a somewhat clearer uniaxial figure may be observed.

From all this may be seen that here we are facing the result of secondary alterations of another primary mineral which may be only biotite, forming a constituent of local gneiss-granites.

The process of biotite transformation during weathering, which is accompanied by weathering of feldspar to be found together with biotite and muscovite, is accompanied in its initial stage by a removal of ferric oxides and—to a lesser extent—of magnesia. Owing to this and possibly also to the presence of a potassium salt which is a product of feldspar weathering, an enrichment in potassium takes place, as it is shown by a total chemical analysis. The weathering conditions were in this case of a reducing character which may be seen from the character of the ammoniac extraction from the

portion of the gneiss-granite weathering products containing the green mineral. Upon treatment of same with a NH_3 solution a liquid has been coloured brownish-yellow owing to the solved ammonium humate.

It is hardly probable that the process of biotite weathering stops here. As has been mentioned above, the Glukhovetz kaolin deposit at the Ukraine shows that not only feldspar but micas too (biotite and muscovite) are entirely passing to kaolinite under adequate climatic conditions.

The intermediate stage of the transformation of biotite into kaolinite, closer approaching kaolinite, is the *x*-mineral of D. S. Beliankin, contained in considerable quantities in the refractory clays of the Chassov-Yar deposit.

The constant retention of Mg and K in the residual weathering product is in no way contradictory to the capacity common to the above elements to alkalize and to interchange with other cations with more difficulty than, say, Na or Ca.

Thus, the character of the Chassov-Yar clay deposit, possessing such a peculiar and enigmatic chemical composition, becomes perfectly clear if regarded as a redeposition of a finely elutriated product of a certain stage of weathering of the micaceous component of gneisses and gneiss-granites. It seems almost beyond doubt that a certain amount of kaolinite is admixed with the given weathering product and, on the other hand, in the widespread kaolinite deposits the predominant constituent, kaolinite, is mixed with different quantities of the *x*-mineral. Indications of this kind may be found in different papers, in which the description of refractory clay deposits is accompanied by a description of their mineralogical composition.

Received
19.VI.1938.

REFERENCES

[1] П. А. Земятченский, Глины СССР, ч. 2, Описание месторождений (1937).

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BIOLOGY

ON THE INFLUENCE OF LOW TEMPERATURES IN FISH

By G. V. SAMOKHVALOVA

(Communicated by A. A. Richter, Member of the Academy, 22.VI.1938)

In the course of breeding *Gambusia* for cold-resistance our attention was largely devoted to the selection for this character in *Cyprinodontidae viviparae*. The data obtained, aside from the purely theoretical interest they present, enable us to draw some practical conclusions in connection with the work of Schmidt and Platonov who were applying anabiosis, induced by low temperatures, to the transportation of live fish. The first experiments carried out in 1935 showed that fishes, when transferred into a vessel in which low temperature was maintained, fell into a state of shock in case the difference in water temperature exceeded a definite value. This condition is characterized by the following: after a short period of excitation the fish ceases all respiratory and swimming movements; it loses the sense of direction, turns on one side or upside-down and gives the full impression of being dead. Replaced in a vessel with water of a higher temperature the fish in a few seconds recovers its normal functions.

In that which followed we succeeded to find out that the temperature of falling into shock might to some extent serve as a criterion of cold-resistance in the species under investigation. Those fishes which fall into a state of shock at lower temperatures usually prove to stand out prolonged falls of temperature for a longer period of time and, on being replaced under normal conditions, quicker recover their functions, i. e. turn out to be cold-resistant. Proceeding from this fact, we started an investigation of cold-resistance in fish, as dependent on species, age, sex and temperature conditions of the surrounding medium. Refrigerating was achieved by placing the vessel with fishes into a crystallizer containing ice. Water temperature was kept under constant control; temperature and rate of collapsing were registered by aid of a stop-watch.

By a series of experiments we succeeded in proving that cold-resistance in fish is dependent on the temperature the fish previously lived at. In order to clear up this question, we studied the temperatures of collapsing in adult fishes, brought up at various temperatures of the aquarium (Table 1).

The data tabulated below point to the existence of a definite dependence between the temperature of the aquarium the fishes were kept in, and the temperature at which they fall into shock; a lower temperature of the aquarium corresponds to a lower shock temperature and, consequently, to a higher cold-resistance of the fish, and vice versa.

To make clear the question whether cold-resistance in fish can be changed experimentally, we undertook special experiments that consisted in transferring fishes from lower temperature to a higher one and vice versa (Table 2).

Table 1

Temperature of Collapsing as Dependent on the Aquarium Temperature

Species	№ of experiment	Number of fishes and sex	Aquarium temperature °C	Average $t°$ of collapsing in °C
Gambusia	1	21 ♀♀	30	9.5
	2	10 ♀♀	28	8
	3	5 ♀♀	26	8—7
	4	21 ♀♀	25	7
	5	16 ♀♀	24	5.5—5
	6	10 ♀♀	21	4.5
Swordfish and hybrids of swordfish with *Platipecilia*	1	10 ♂♂ and 2 ♀♀	29	14—13.5
	2	10 ♂♂ and 3 ♀♀	27	13.25—12
	3	10 ♂♂ and 3 ♀♀	26	13—12.5
	4	3 ♂♂	25	12
	5	10 ♂♂ and 10 ♀♀	24—25	11.5—11
	6	10 ♂♂	22	10—9
	7	6 ♂♂	21	10
	8	4 ♂♂	20—21	9
	9	3 ♀♀	18	8

Table 2

Character of experiment	Number and sex	Temperature at which the shock set in		
Change in shock $t°$ when transferred from higher to lower $t°$	—	Exp. 1 (aquarium $t°$ 26—27 °C	Exp. 2 (2 weeks after the $t°$ had been reduced to 21 °C)	
	—			
	—			
	3 ♀♀	12.5	8.75	
	3 ♀♀	12	10.5	
Change in shock $t°$ when transferred from lower to higher $t°$		Exp. 1 aquarium $t°$ 13 °C)	Exp. 2 (10 days after the aquar. $t°$ had been raised to 21.5°)	Exp. 3 (2 weeks after the aquar. $t°$ had been raised to 27°)
	3 ♀♀	6	10	12.92

In the experiment illustrated by Table 2 the shock temperature has been determined for fishes taken from two different aquariums with a temperature from 26—27° C. After this the temperature of the aquarium was made to fall to 21° C by switching off the heaters, and in two weeks the shock temperature was determined again. The experiment showed that fishes kept at lower temperatures fell in a state of shock at a correspondingly lower temperature, i. e., they became more cold-resistant.

An analogous experiment carried out in the inverse sequence showed that being transferred to higher temperatures, the fishes became less cold-resistant (see 2d half of Table 2).

In Table 3 are summarized comparative data as to the temperatures corresponding to the initial and repeated shock.

The experiments were thus arranged that soon (20—30 minutes) after the first cooling the fishes were subjected to the second cooling.

Table 3

№ of exper.	Number of fishes and sex	Aquarium $t°$ in °C	$t°$ of the initial shock	$t°$ of the repeated shock 30 min. after the initial one
1	3 ♀♀	28	12.7	10.7
2	3 ♀♀	18	9	8
3	4 ♂♂	23.5	11	9.25
4	4 ♂♂	27	12.63	10.53
5	1 ♀ and 3 ♂♂	26.5	13.81	10.5
6	1 ♀ and 2 ♂♂	27	12.17	10
7	4 ♂♂	22	10.5	9
Total . . .	8 ♀♀ 17 ♂♂	From 18—28°	11.70	9.68

From Table 3 may be seen that the fishes get rapidly accustomed to the falls of temperature, collapsing at repeated coolings at lower temperatures, which points to their becoming more cold-resistant. The second cooling, performed 30 minutes after the first one, results in a fall of the shock temperature averaging 2°.

All the experiments described thus show that cold-resistance in fish depends on the temperature they have been kept at prior to the experiment. When transferred from one temperature to another, their cold-resistance is correspondingly modified, growing with the reduction of the temperature of the medium, and decreasing with the rise of the latter.

In a series of experiments we succeeded to find out that cold-resistance in fish is linked with age. Thus, in *Gambusia holbrooki* the young are usually less cold-resistant than adult specimens. Out of 40 females at the age of above 2 years taken at 24°, all fell in a state of shock at the temperature of 5° C; while 50 young fishes at the age of several months taken from the same aquarium at the same temperature, collapsed at about 8° C, i. e. they proved to be less cold-resistant.

The data in Table 4 show that the males of *Gambusia* are less cold-resistant than the females. The fishes were taken from the same aquarium where they were reared under the same conditions. When subjected to cooling, the males were first to fall in a state of shock, which occurred on the average after 6.8 minutes at 10° C, while the females collapsed after 8.5 minutes at 7°. It is noteworthy that analogous experiments made with other species of viviparous fishes showed in swordfish and especially in *Platipecilia* a more pronounced sex variation with regard to cold-resistance.

Of great interest and importance is the study of specific cold-resistance in fish. The data given in Table 1 are sufficient to show that *Gambusia* is more cold-resistant as compared with the swordfish and with the hybrids of the swordfish with *Platipecilia*, since *Gambusia* taken at the same temperature falls in a state of shock at lower temperature. The variation in the shock temperature in both species is strictly constant and equalizes about 5° C (Table 1).

In the subsequent experiments made with the purpose to investigate specific variation of cold-resistance in fish, we were comparing *Gambusia* with *Phallocerus caudomaculatus* and *Lebistes reticulatus* respectively.

The fishes were taken from aquariums with the same temperature; they were of the same sex and as far as it was possible under the experimental conditions, of the same size.

The cooling of the fishes of the two species to be compared was carried out simultaneously. The experiment has shown that the females of *Phallocerus caudomaculatus* collapsed first, 3.77 minutes after the beginning of the experiment, at 13.75° C, while all the females of *Gambusia* fell

Table 4

Sex Varition of Shock Temperature in *Gambusia*

Sex	Dimensions		Shock $t°$ in °C	Rate of collapsing time since beginn. (of exper. in min.)
	length in mm	width in mm		
♂	29	4.5	12	6.1
♂	26	4.5	11	6.2
♂	26	4	11	6.6
♂	29	5	10.5	6.7
♂	30	5	9	7.8
♀	40	9	7	8.3
♀	44	10	7	8.35
♀	45	10	7	8.5
♀	46	10	7	8.6
♀	47	10	6.75	8.7
5 ♂♂	28	4.6	10.75	6.68
5 ♀♀	44.4	9.8	6.95	8.41

in the state of shock the last, at 11°. In another analogous experiment carried out on males, the males of *Phallocerus caudomaculatus* collapsed first as compared with the males of *Gambusia*. The same results were obtained in comparing *Lebistes* to *Gambusia* as to their respective cold-resistance. Five females of *Lebistes* fell into a state of shock first, at 10.3° C, five females of *Gambusia* previously kept under the same conditions, were the last to collapse, at 5° C. All these experiments have shown that *Gambusia* is more cold-resistant as compared to both other species mentioned above.

Thus, the data given above point to a definite degree of cold-resistance inherent to each species of fish.

P. Schmidt and G. Platonov (2) have observed variation in cold-resistance in fishes caught in winter and in summer respectively, which these authors referred to the seasonal response of the fish to the action of low temperatures.

In the light of data obtained by us, the seasonal character of the response in the given case seems to be accountable for by the difference in water temperature, to which the fish is subjected during winter or summer.

The fish caught in autumn or winter and taken out of water with low temperature, must be more cold-resistant than that caught in summer.

Experiments with the transferring of fishes from one temperature to another have shown them to possess a rather developed adaptability to temperature variations. These experiments permit of a rather important practical conclusion, namely, that subjecting the fish to a preliminary treatment by keeping it for some time at low temperature, its cold-resistance can be raised. On the other hand, by making fishes more cold-resistant, the latter can be made apt to transportation in a refrigerated state at any season. The duration of the preliminary treatment neccessary seems to be specific.

It is quite obvious that each species of commercial fish should be studied as to its cold-resistance and adaptability to falls of temperature, with special reference to age and sex. The last two factors should also be taken into account in selecting material fit to be transported in live state.

Thus, from the data obtained we have come to the conclusion that cold-resistance in fish, as determined by shock temperature, is dependent on species, sex and age, as well as, to a considerable degree, on the water temperature at which the species under investigation had been kept previous to the experiment.

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REFERENCES

[1] P. Schmidt et G. Platonov, C. R. Acad. Sci. URSS, № 5 (1937).
[2] P. Schmidt et G. Platonov, ibid., № 3 (1938).

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MICROBIOLOGY

THE ALTERATIONS OF CHONDRIOSOMES OF YEAST ORGANISMS IN THE COURSE OF RESPIRATION AND FERMENTATION

By M. N. MEISSEL

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Chondriosomes have been recognized as one of the necessary cellular organoids not so much on the ground of our understanding the part played by them in the organism, as on that of the remarkable constancy of their occurence in the protoplasm of almost all the known cells.

Most of the studies, dealing with the elucidation of the rôle of chondriosomes, show two essential defects: they either ascribe to the chondriosomes almost the entire complex of functions inherent to the cell, leaving to the cytoplasm itself an insignificant part only, or attempt to ascribe to the chondriosomes of most various cells one and the same particular and constant function.

These views originate from the notion of the chemical composition of chondriosomes as being similar and constant.

However, recently a number of facts have accumulated, pointing to a different chemical composition of chondriosomes of different cells. Particularly important is the fact that such highly active substances, as glutation, vitamins and various enzymes, are present in the chondriosomes of different cells in various combinations.

There are grounds to believe that chondriosomes are structural adjustments, aiding in the concentration (or may be in the respective spatial orientation) of highly active catalytic substances, the character of the latter depending on the kind, the specific differentiation and the state of the life-activity of the cell.

The idea of the enzymatic function of chondriosomes taken in a wider sense finds of late an ever increasing number of adherents [17].

Joyet-Lavergne [8,9] is doing much work for the elucidation of the eventual rôle of chondriosomes in the oxidation-reduction processes.

However, the overwhelming majority of the data obtained by this author must be regarded as indirect evidence only. In his recently published work Joyet-Lavergne [10] makes an attempt to establish a relation between the number of chondriosomes in the cells of the triton and the intensity of respiration of these cells.

Unicellular organisms, which allow an intravital cytological observation and a simultaneous accurate registration of their physiologic responses, may be particularly valuable for further studies, aiming at the elucidation of the rôle of chondriosomes.

Yeasts and yeast-like organisms proved quite convenient objects for studying chondriosomes under various functional conditions of the cell. In an previous study ([11]) I noted the typical pictures of chondriosomes in yeasts which corresponded to aerobiotic respiration and to alcoholic fermentation.

In the first case, i. e. under conditions favouring aerobiotic respiration, the chondriome of yeast cells is composed by numerous fine chondriosomes which are distributed more or less regularly throughout the cytoplasm. Under the conditions of fermentation a sharp alteration is observed in the chondriosomes. It is expressed in their fusion, hypertrophy, migration towards the surface of the cell and in the appearance of signs of a peculiar «secretion» on them. In special experiments we succeeded to establish that this reconstruction of chondriosomes is intimately connected with the active participation of the cell in the process of fermentation and does not depend on the accumulation of metabolic products in the medium or on any other unfavourable condition.

Subsequent research has shown that a reconstruction of the chondriosomes, similar to the one we observed in fermentation in a liquid sugar containing medium, can be obtained likewise on a solid medium (e. g. on wash-agar) by subjecting the culture to conditions of a relative vacuum or of a decreased oxygen content. In these cases, similarly to what we saw in fermentation in a liquid medium, a sharp hypertrophy of individual chondriosomes and their fusion take place. The cytologic picture of yeast organisms developing under anaerobiotic conditions resembles greatly the one observed in active fermentation (accumulation of metachromatin within the vacuoles, increase of the size of the nuclei, intravital staining of the latter, etc.). In the absence of fermentable carbohydrates or at low temperatures, which depress fermentation, the chondriome of the cells preserves even in a liquid medium the type which is characteristic of breathing yeast cells. This shows that the character of the chondriosome's structure depends on the conditions of the life-activity of yeast cells and is closely bound with the prevailing type of metabolism. We were able to detect with perfect clearness the connection between the condition of the chondriosomes and the respiratory and fermenting functions of the cells in the same culture of yeast organisms in the course of their cytological study, done parallel with the registration of their respiration and fermentation in a Warburg apparatus. It was found thats yeast of the same strain, but developing either under conditions favouring respiration, or under those of fermentation, acquire not only the cytological structure typical of each of these conditions, but the respective physiological peculiarities, corresponding to these conditions, as well. Thus, the yeast cells *Saccharomycodes Ludwigii* which had developed during 48 hours on a solid medium (wash-agar) breathe 2.5 times more intensely than the cells of the same strain, grown in a similar but liquid medium (beer wash), and ferment less than half the amount of glucose, fermented by the latter.

These peculiarities are preserved by yeasts for several hours after they have been washed in sterile water (or phosphate buffer) and subjected to an identical subsequent treatment of the samples, used for respiration and fermentation tests. The parallel study of the cytological structure of the cells shows a complete coincidence of the structure of chondriosomes with the corresponding physiological peculiarities noted above.

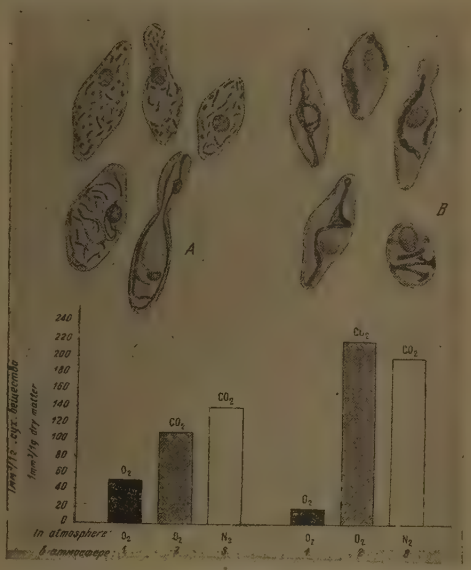
The table shows the cells of *Saccharomycodes Ludwigii* both from the solid medium (conditions favouring respiration) and from the liquid one (conditions favouring fermentation), and supplies the respective graphs, which illustrate the intensity of respiration and fermentation of these cultures after their having been washed thrice and tested in 2% solution of glucose.

Similar data have been obtained by us in studying cultures which had developed on solid media under conditions of a relative vacuum or of oxygen deficiency. In these cases the yeasts corresponded both cytologically and physiologically to cultures in a state of fermentation.

The morphological reconstruction of the chondriosomes taking place when the cells pass from breathing to fermentation is accompanied also by their functional alterations.

Thus, intravital staining tests (Janusgrün) demonstrate that the ability of chondriosomes to absorb the stain and to reduce it in the cells, in which breathing prevails, is different from the one observed in the cells with prevailing fermentation. The latter show an increased ability to absorb Janusgrün, and as a result of this their chondriosomes stain more completely and intensely.

Yeast, cultivated in media, containing considerable concentrations of substances which depress either respiration or fermentation (potassium cyanate, ethylic alcohol, chloralhydrate, mono-iodine-acetic acid), shows a reduced chondriome. Individual chondriosomes become fine, thin, sometimes are dispersed like dust throughout the cytoplasm. Attempts to stain the chondriome intravitally in such cells usually fail, even after a thorough washing, or else the chondriosomes stain so faintly that they remain hardly discernible.



A — chondriome of *Saccharomycodes Ludwigii* which developed under good aeration conditions: *B*—chondriome of *Saccharomycodes Ludwigii* under conditions of fermentation. Under these is a graph showing the values of respiration (*1*), aerobiotic (*2*) and anaerobiotic fermentation for the respective cultures (*A* and *B*).

Under the action of narcotic substances (alcohol, ether, chloroformium) even in relatively weak concentrations, when no pronounced morphological alterations of the chondriosomes are as yet observed, the ability of the chondriosomes to adsorb and concentrate at their surface the Janusgrün stain is depressed and even suppressed.

After the elimination of the narcotics the staining of the chondriosomes in these cells is restored.

The same loss of staining of the intravitally stained chondriosomes occurs also if the cells are transferred into a liquid medium (beer wash). Uninterrupted observation of such preliminarily stained cells shows that $1—1^1/_2$ hours later the chondriosomes gradually lose their staining, and after this the yeast cell acquires the ability to participate in fermentation and, in the following, that of multiplication. Cells in which the surface of their chondriosomes is blocked by intravital staining are incapable of fermentation and they must free themselves of the stain, before they can completely resume their functioning.

The liberation of the chondriosomes of a live cell from the stain can proceed, as the experiments have shown, along two ways: either through

an increase in the processes of reduction, as a result of which the greenish-blue Janusgrün is transformed first into a pink and then a colourless leuko-form, or through a decrease of the adsorption capacity of the chondriosomes, which leads to the displacement of the stain from the surface of the chondriosomes onto the surrounding cytoplasm. It was found that under different conditions of the cell's life-activity sometimes one sometimes the other of these mechanisms may be operative.

All these data allow us to draw the following conclusions:

1. The chondriome of yeast cells suffers an essential morphological reconstruction when these organisms pass from conditions favouring respiration to those of fermentation.

2. This reconstruction coincides fully in time with the change of the relation between respiration and fermentation, which is observed when yeast organisms turn from an aerobiotic to an anaerobiotic existence. Thus, one can on sound grounds distinguish in yeast organisms the chondriome type, corresponding to the aerobiotic, and the one corresponding to anaerobiotic metabolism.

3. Factors which depress respiration or fermentation affect the chondriome, inducing at first a change in its functional properties (ability to reduce, oxidize and accumulate stains) and in case of a more considerable action—in its morphological structure as well.

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1. VII. 1938.

REFERENCES

[1] Robertson, Austr. Journ. Exper. Biol. a. Medic., **3**, 97 (1926). [2] Marston, ibid., 233 (1926). [3] Cowdry, Amer. Natur., **60**, № 667, (1926). [4] Horning, Austr. Journ., Exper. Biol. a. Medic., **3** (1936). [5] Horning a. Petri, Proc. Roy. Soc., ser. B., **102**, 188 (1927). [6] Bourne, Austr. Journ. Exper. Biol. a. Medic., **8**, 239 (1935). [7] Smith, Journ. Morph. and Physiol., **52**, 485 (1931). [8] Joyet-Lavergne, La Cellule, **43** (1934). [9] Joyet-Lavergne, C. R., **203**, 1020 (1936). [10] Joyet-Lavergne, ibid. **203**, 1291 (1936). [11] Мейсель, Бюлл. экспер. биол. и мед., **5** (1938).

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GENETICS

DEPENDENCE OF THE DEVELOPMENT OF THE WHEAT PLANT ON THE GEOGRAPHIC ORIGIN OF SEED

By N. V. SEMIGANOVSKAIA

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It was noted by Kostiuchenko in 1935 at the Polar Experiment Station of the All-Union Institute of Plant Industry, Khibiny (67°44′ n. l.) that winter wheat plants grown from seed produced at Kirovobad (40°41′ n. l.) strikingly differ from plants of the same variety grown from seed produced at Khibiny. The difference consisted in the fact that the seed produced at Khibiny when sown in spring without preceding artificial yarovization gave plants which developed ears and partly ripened, while the seed from Kirovobad which was sown simultaneously gave plants that did not develop beyond the tillering phase. This behaviour of plants was explained as the result of yarovization produced by low temperatures upon the maturing seeds; a series of experiments were arranged to verify this explanation [1,2,3].

Taking into consideration the importance of these data on the significance of the origin of seeds for the subsequent development of plants grown there from the Section of the Georgaphic Cultures of the All-Union Institute of Plant Industry (VIR) sowed in 1937 at Khibiny samples of wheat belonging to various ecological-geographic groups and possessing various degrees of winterness in order to compare one and the same form raised from seed produced at different localities. Seed of winter wheat and half-winter wheat produced from winter cultures (without yarovization) at Tashkent (41°20′ n. l.) and that of the local cultures raised at Khibiny in 1936 from non-yarovized seed sown in spring were used for the experiments. Sowing was done on the 20 May in the field on plots with yarovized and non-yarovized seed. Since natural yarovization is taking place at Khibiny as a result of low spring temperatures, the duration of the preliminary yarovization was reduced. Forms with a long yarovization stage (from Holland and Denmark) were yarovized during 58 days, those with a short one (from Japan and China) for 43 days, and the remainder—during 48 days.

Conditions of the vernal yarovization in 1937 were more favourable than in 1936. In 1936 the mean temperature of the soil at the depth of 5 cm was 6.8° in the III decade of May, 13.7° in the I decade of June; in 1937 the temperatures were 6.4° and 10.1° respectively. Owing to these lower temperatures in 1937 the stimulation of earing after artifical yarovization was less than in 1936 (see the table, columns 7 and 10). The conditions of maturation were also considerably more favourable in 1937 than in 1936; a great number of forms attained full waxy ripeness in the first days of September, those which did not mature were harvested on the 14th of September. The results of the experiment are presented in the table.

№ of the catalogue of VIR	Origin of the form and variety	Locality of seed grown	1936			1937			Weight of seeds from 10 plants (in g)		Weight of 1 000 seeds (in g)	
			Phase of harvest or number of days till waxy ripeness		Acceleration of earing after artificial yarovization (days)	Phase of harvest or number of days till waxy ripeness		Acceleration of earing after artificial yarovization (days)				
			non-yarovized	yarovized		non-yarovized	yarovized		non-yarovized	yarovized	non-yarovized	yarovized
1st group (close to summer wheat)												
28 613	North China, Shan-Si, *splendens*	Tashkent	Milk ripeness	Close of milk ripeness	5	83	86	—4	7.8	No data	32.0	No data
		Khibiny	—	—	—	81	84	—1	4.6	9.8	33.4	33.3
19 389	Poland, Lwow, *ferrugineum*	Tashkent	Beginning of waxy ripeness	90	4	85	87	—2	12.5	11.4	24.2	25.1
		Khibiny	—	—	—	84	84	0	11.6	17.1	27.2	28.1
26 426	USA, *lutescens* «Boort×Ridit»	Tashkent	Milk ripeness	Close of milk ripeness	8	89	84	5	12.0	15.9	50.4	48.1
		Khibiny	—	—	—	83	85	—1	7.2	6.3	40.9	39.0
2nd group (half-winter wheat)												
28 589	North China, Shan-Si, *albinflatum*	Tashkent	Beginning of grain formation	91	40	Milk ripeness	87	13	7.8	14.0	28.3	35.1
		Khibiny	—	—	—	Beginning of waxy ripeness	82	8	15.3	15.2	31.0	31.4

6 018	France, *albidum*	Tashkent	Flowering	93	46	Milk ripeness	84	18	5.4	8.8	40.6	44.4
	»Richelle blanche native«	Khibiny	—	—	—		85	11	10.9	11.5	46.6	45.2
26 721	USA, Minnesota *erythrospermum* «Minhardi»	Tashkent	Earing	End of milk ripeness	42	Earing	81	33	—	15.3	—	30.1
		Khibiny		—	—	Beginning of waxy ripeness	83	6	5.8	12.6	30.6	27.3
						3rd group (winter wheat)						
26 256	USA, *erythrospermum*	Tashkent	Tillering	Beginning of waxy ripeness	∞*	Beginning of earing	85	45	—	11.8	—	39.2
	«Kanred»	Khibiny	—	—	—	Beginning of waxy ripeness	87	16	5.9	5.8	32.7	38.6
27 068	Czecoslovakia. *milturum* «Sudmaricher Winterweizen»	Tashkent	Tillering	92	∞	End of flowering	93	36	—	No data	—	36.5
		Khibiny		—	—	Beginning of waxy ripeness	89	14	7.3	10.2	27.4	No data
22 984	Yugoslavia, *erythrospermum* «Sortovaia»	Tashkent	Tillering	92	∞	Beginning of tillering	87	47	—	21.5	—	46.2
		Khibiny		—	—	Milk ripeness	86	19	7.1	22.0	34.0	44.5
17 517	Armenia, *erythrospermum*	Tashkent	Tillering	End of milk ripeness	∞	Beginning of earing	93	50	—	14.4	—	44.8
		Khibiny			—	Beginning of waxy ripeness	93	10	No data	19.4	No data	47.6
11 389	Ukr. SSR, *erythrospermum* «Kooperatorka»	Tashkent	Tillering	Waxy ripeness	∞	Beginning of earing	86	47	—	17.7	—	38.8
		Khibiny	—	—	—	Beginning of grain formation	89	36	0.6	16.2	—	39.0

№ of the catalogue of VIR	Origin of the form and variety	Locality of seed grown	1936			1937			Weight of seeds from 10 plants (in g)		Weight of 1 000 seeds (in g)	
			Phase of harvest or number of days till waxy ripeness		Acceleration of earing after artificial yarovization (days)	Phase of harvest or number of days till waxy ripeness		Acceleration of earing after artificial yarovization (days)				
			non-yarovized	yarovized		non-yarovized	yarovized		non-yarovized	yarovized	non-yarovized	yarovized
4th group (winter wheat)												
25 125	Denmark, *albidum*	Tashkent	Tillering	Beginning of waxy ripeness	∞*	Tillering	Beginning of waxy ripeness	∞	—	21.7	—	39.6
	«Trifolium wheat»	Khibiny	—	—	—	Beginning of grain formation	Ditto	12	2.5	17.0	—	36.1
25 150	Holland, *albidum* «Emma»	Tashkent	Tillering	Milk ripeness	∞	Tillering	Beginning of waxy ripeness	∞	—	15.7	—	36.5
		Khibiny	—	—	—	Flowering	Ditto	20	—	14.7	—	32.7
2 704	Czekoslovakia, *ferrugineum* «Winterweizen»	Tashkent	Tillering	92	∞	Tillering	89	∞	—	14.3	—	35.0
		Khibiuy	—	—	—	Flowering	86	34	—	11.0	—	37.3
5th group (winter wheat)												
25 519	Japan	Tashkent	Tillering	Milk ripeness	∞	Tillering	94	∞	—	16.5	—	40.8
		Khibiny	—		—	Tillering	88	∞	—	No data	—	41.1

* The sign ∞ signifies that the sample did not reach the earing phase and the acceleration of the earing after artificial jarovization can consequently not be expressed.

For the sake of convenience the samples are divided in the table into five groups; the behaviour of each of them at Khibiny is thus characterized separately.

1. Into the first group are included those wheats which but slightly react to the artificial yarovization. When grown from non-yarovized seed of local

Left, winter wheat «Emma» from Holland, right, *erythrospermum* 17517 from Armenia; *Ia* and *IIa*, plants from seeds produced at Khibiny; *Ib* and *IIb*, plants from seed produced at Tashkent.

origin they show insignificant reduction of the growth period as compared with those raised from southern seed. At other localities where the geographic cultures were tried these samples behave as summer forms.

2. To the second group are referred wheats whose response to artificial yarovization is well pronounced. The growth period is considerably shortened when non-yarovized seed of local origin is used in contrast to plants obtained from southern seed. At other localities these wheats behave sometimes as winter forms, sometimes as half-winter forms, depending upon the temperature conditions of the spring.

3. Into the third group are included those wheats in which earing is significantly accelerated by artificial yarovization. These forms, when raised from southern seed, behave sometimes as winter forms, sometimes as half-winter forms, depending upon the spring temperature; however, they fail to attain the stage of grain production. On the other hand, plants raised

from local seed produce grains of various degrees of maturity, from the beginning of grain formation up to the beginning of waxy ripeness. Both this group and the following ones behave at other points of geographic cultures as typical winter forms.

4. The wheats referred to the fourth group, when raised from southern seed, remain at the stage of tillering throughout the growth period. When raised from local seed the same varieties produce plants which enter the phase of flowering at the end of the period of growth and even attain grain formation; under the conditions of warm spring, however, they need 70 days of artificial yarovization for earing the same year.

5. Finally, into the last group may be classified only one sample from Japan which remained at the tillering stage even when raised from the local stage. This variety has a comparatively short yarovization stage, it is earing and ripening under the conditions of Khibiny (when raised from seeds artificially yarovized during 43 or even 37 days). However, it seems to demand for passing the yarovizatoin stage stable, although not long, low temperatures which cannot be obtained in spring under natural conditions.

The data obtained enable us to note the following:

1. The experiment conducted confirms the possibility of a totally different behaviour of one and the same variety (belonging to winter or to half-winter wheat), when raised from seed produced by plants grown under different climatic conditions.

2. Plants of winter and half-winter wheats obtained from the spring sowing of yarovized seed at Khibiny produce progeny which behave, when raised the following year from non-yarovized seed, in the following manner: forms closely related to summer wheat somewhat shorten their period of growth; half-winter forms reach full maturity or the beginning of waxy ripeness; winter forms, depending on the duration of the yarovization stage, give grain of various stages of maturity or only reach the flowering stage. Cases may occur when field conditions do not satisfy requirements necessary for passing the yarovization stage, and a given form may stay at the tillering stage (in our experiment there was one such case out of 15).

3. The longer the yarovization stage, the more expressed is the difference in the rate of development in plants raised from seed produced at Tashkent and at Khibiny.

4. The difference in the rate of development of plants raised from seed of different origin nearly disappears when seeds are artificially yarovized.

Whether the acceleration of the development of plants raised from local seed is the result of yarovization of seed while on the mother plant under the influence of the low temperature of the autumn, or it is the result of the summarized action of the climatic factors of the North, those are questions to which our one-year experiment fails as yet to give any answer. In our further work we shall try to obtain a more exact information as to the degree and the character of alterations in the development and growth of the wheat belonging to different agro-ecological types, when subjected to the influence of cultivation under the conditions of Extreme North.

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REFERENCES

[1] И. А. Костюченко и Т. Я. Зарубайло, Селекция и семеноводство, № 3/11, 49—53 (1935). [2] И. А. Костюченко и Т. Я. Зарубайло, там же, № 6, 39—42 (1937). [3] F. G. Gregory a. O. N. Purvis, Annals of Botany, II, № 5, 237—251 (1938).

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ZOOLOGY

ON THE QUESTION OF GLACIAL-MARINE RELICTS IN FRESH WATER FAUNA

By S. S. SMIRNOV

(Communicated by S. A. Zernov, Member of the Academy, 5.VI.1938)

In 1927 V. A. Jashnov(1) described after young specimens from the northern part of the gulf of Ob, a form of *Copepoda* which he nominated *Gaidius* sp. 2. When working on the material collected in the gulf of Ob and the Gyda bay by the Ob-Tazov Expedition of the All Union Arctic Institute, I discovered this form in prodigious numbers, and moreover, in addition to young specimens, I came across a considerable amount of adult females. However, in order to accurately determine the taxonomic position of this form, it proved necessary to study adult males which were absent in that material, though the number of males of the last larval stage were very considerable. E. V. Burmakin, who had worked in the Gyda bay in 1937 and collected a great deal of material on plankton over there, kindly put to my disposal a number of specimens of said form, which included one quite mature male. Jashnov's investigation of adult specimens of *Gaidius* sp. 2 has shown that this form may be identified with *Senecella calanoides* Juday, described from the lakes of North America and as yet unknown in Eurasia.

A comparison of Siberian specimens of *Senecella calanoides* with those described in literature (4,5,6) showed a great similarity of our specimens with the North American, except that some specimens from the gulf of Ob are considerably larger than the American ones. Figures showing the fifth pair of feet of the male *Senecella* from the Gyda bay, are somewhat different from those given by the American authors, whom I have quoted, however, not to such a degree as to give vent to question the specific position of our specimens. I should still mention that the fourth pair of feet, at the inner angle of the first segment of the basipodite, bear an appendage made up of an essential solid part and a fine seta, the latter forming the distal half of the

appendage. Some specimens had the distal half broken off. It seems to me that the specimens Juday and March dealt with had also had the distal half of the appendage broken off.

Senecella calanoides was recognized by me in the northern part of the gulf of Ob, beginning approximately from the parallel at Mys Shtormovoy, i. e. in the zone of waters with salinity over 1‰. But this form, probably, thrives also in the more southerly parts of the gulf of Ob, with lower salinity (whence there was no material in my possession), for in the Gyda bay it reaches the furthermost south-eastern corner of the bay, where only insignificant quantities of chlorine have been found in the water. As to the maximum salinity the form under investigation is able to withstand, there is no limit we can set, for all we know is that *Senecella* occurs in the waters of the south-eastern part of the Kara Sea with a somewhat lower salinity than the ordinary marine salinity. There is no doubt whatever that *Senecella calanoides* belongs to forms quite euryhaline, however, data available at the present time point to this form reaching its utmost development after all under conditions of strong admixture of fresh water (salinity not exceeding 15—20‰).

The size of *Senecella calanoides* from northern Siberia is subject to considerable variation. The following table presents measurements (in mm), taken of specimens from various stations.

	Males				Females					
	Adults	Fifth Copepodan stage			Adults			Fifth Copepodan stage		
	Size	Number of specimens	Range of variation	Average size	Number of specimens	Range of variation	Average size	Number of specimens	Range of variation	Average size
Ob-Tazov Expedition of All-Union Arctic Institute:										
Northern part of the gulf of Ob, by the village Taran, 26.IX. 1934	—	15	2.50—2.65	2.59	8	3.15—3.35	3.25	15	2.73—2.91	2.70
E. V. Burmakin's collections										
Gyda bay, st. 55, 15.IX. 1937 . .	—	30	2.23—2.61	2.45	1	2.95	—	13	2.33—2.73	2.55
Gyda bay, st. 53, 15.IX. 1937 . .	—	9	2.28—2.49	2.38	2	2.90—2.93	—	9	2.31—2.69	2.41
Gyda bay, st. 25/26, 25.IV. 1937 .	2.36	—	—	—	2	2.62—2.78	—	—	—	—
Lake Yambu-To, st. 1, 3.VIII. 1937	—	—	—	—	—	—	—	2	1.98—2.06	—

From these scanty data may be seen that *Senecella Calanoides* diminishes in size in the direction from the sea towards the mouth of the river Gyda, this being manifest not only in the sizes of mature animals, but in the sizes of larvae (the last Copepodan stage) as well. It should be remembered, that

Senecella calanoides from North American lakes have measurements most closely corresponding with sizes of the specimens from the station 25/26 of the Gyda bay. It is quite possible, though from available data it could not be stated with absolute certainty, that the sizes of *Senecella* depend upon the degree of salinity of the water they inhabit. We feel more certain as to the term of their attaining maturity. Mature females *Senecella* are met only beginning from the middle of September; on the other hand, the April catch by E. V. Bumarkin yielded one mature male and two mature females *Senecella*. Maturity seems to be reached during the second half of September (in males apparently even later than that), and mature animals thus hibernate. In the summer, on the other hand, none except Copepodan larvae of *Senecella* are met. In this respect, Siberian aggregations of *Senecella* are entirely different from North American, as in North America mature males and females were caught in July and August. Unfortunately, the material from the lake Yambu-To was too small (there were altogether three larvae of *Senecella*) to contribute anything to our knowledge on the subject. I should note that the mouth parts in young males (of the 5th stage) *Senecella* investigated by me were well developed and did not differ from those of the females, whereas the only one mature male which I had at my disposal exhibited quite strongly reduced mouth extremities, such observation having already been made by American authors. Evidently here we come across the same phenomenon as in the case of *Calanus tonsus* Brady [after the material of Campbell(2)].

In North America *Senecella calanoides* has been found in the following lakes [Juday (5), Marsh (6)]: Seneca Cayuga, Owasco (New York); Pine Lake (Michigan); the Upper Lake near Duluth; Timagani Nipigon (Canada). In his letter of April 21, 1938, Prof. Ch. Juday kindly informed me that this species was found also in the «Lake Champlain and another in the State of New York». In North Siberia the same species was found in the gulf of Ob, the Gyda bay, in brackish waters of the Kara Sea. V. L. Khmyznikova found *Senecella* in the Vilkitzky Strait by the island «Komsomolskaya Pravda» (form. Samuel island), in the Br. Laptev Sea and in the brackish water region of the sea by the delta of the Lena river. And last, *Senecella* was also discovered in the fresh water lake Yambu-To lying approximately 150 km to the south-east of the mouth of the Gyda river.

The taxonomic position of *Senecella* is estimated by various authors in different ways. Juday simply points out its resemblance to the marine *Calanoida*, Gurney (3) refers it to the family *Pseudocalanoidae*, and finally Marsh isolates *Senecella* into a separate family without giving a description of the latter or grounding the necessity of such classification. The close relationship of *Senecella* to genuine marine forms is absolutely undisputable.

This circumstance coupled with the peculiar distribution of *Senecella* induce one to dwell upon the question as to how and when such a mode of distribution had been established. Involuntarily a comparison suggests itself with the genus *Limnocalanus* whose region of distribution may comprise the total areal of *Senecella*. Bearing in mind this fact, as well as the ecological peculiarities of *Senecella* in Northern Siberia, we may consider quite possible that the history of penetration of *Senecella* into fresh waters both in America and in Siberia coincides in the main with the history of the migration of *Limnocalanus*, although there is no complete similarity between the two. Thus, the time of penetration of *Senecella* from northern seas into fresh waters should be referred by us to the Pleistocene epoch, when a whole series of animals from those parts of the Arctic Ocean which have fresh water admixture had either actively moved into the fresh waters or were carried into relict water basins which had remained after the boreal sea transgres-

sion. The lake Yambu-To lies in the region of this transgression, but as to the North American lakes, most of them had also been covered by sea during the boreal transgression, while other lakes lay nearby, and the penetration of *Senecella* thither may be easily explained, especially since they are inhabited by *Limnocalanus macrurus* too.

The similarity in ecology and distribution of *Senecella* and *Limnocalanus*, as well as the fact that the penetration of these genera into fresh water seems to have an analogous history, lead to the suggestion that the *Senecella* (as it is known in regard to *Limnocalanus*) possibly includes a fresh water form, different from the original form living in brackish waters. This is, perhaps, to a certain extent confirmed by the above-mentioned data on the difference in sizes of *Senecella*. However, the material at my disposal at the present time would not constitute a sufficient basis, and if studied would not enables one to answer the question raised, the more that to solve the problem it would be absolutely necessary to make a detailed comparison of our specimens of *Senecella* with those from North America.

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REFERENCES

[1] В. А. Яшнов, Тр. Плов. морск. научн. ин-та, II, вып. 2 (1927). [2] M. H. Campbell, Journ. Biol. Board of Canada, I, № 1 (1934). [3] R. Gurney, British Fresh-Water Copepoda, 1 (1931). [4] Ch. Juday, Science, 58 (1923). [5] Ch. Juday, Proc. Un. Stat. Nat., 66, art. 4 (1925). [6] C. D. Marsh, ibid., 82, art. 18, 1933).

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MORPHOLOGIE EXPÉRIMENTALE

RÉACTION THYRÉOTROPE CHEZ L'EMBRYON DE POULE

Par A. N. STOUDITSKY

(Présenté par I. I. Schmalhausen, de l'Académie, le 16.VI.1938)

La réaction de la glande thyroïde à l'hormone dit thyréotrope a été étudiée par Loeb et Basset (8), par Aron (1) et plus tard par beaucoup d'autres auteurs encore. À l'heure actuelle cette réaction est reconnue spécifique et présente des indices morphologiques bien déterminés chez la plupart des vertébrés. L'hormone thyréotrope, secrété par le lobe antérieur de l'hypophyse introduit dans l'organisme, provoque dans la glande thyroïde une série de modifications caractéristiques qui consistent en l'allongement en hauteur de l'épithélium folliculaire, la délimitation marquée des cellules et une dilution de la colloïde. En même temps on observe une forte hypérémie qui aboutit à l'augmentation du volume de la glande.

La réaction thyréotrope a été suivie principalement chez les mammifères jeunes et adultes, ainsi que chez les oiseaux. La réaction de la glande thyroïde à l'hormone thyréotrope chez l'embryon ne peut manquer d'intérêt étant donné qu'elle reflète l'état physiologique de la glande, jouant dans l'autogénèse un rôle qui n'est pas dépourvu d'importance. On connaît les troubles profonds de développement de l'embryon dans l'hyperthyroïdisme et l'athyroïdie embryonnaire [Dieterle (3), Križenecky (6), Hammett (4), Landauer (7), Kraus(5)]. Malgré l'intérêt évident de la réaction thyréotrope de la glande thyroïde chez l'embryon, il est rare de rencontrer dans la littérature des tentatives de déclencher expérimentalement l'hyperthyroïdisme chez l'embryon. Woodside (11) en 1937 a étudié la réaction thyréotrope chez des embryons de poulets en injectant l'extrait du lobe antérieur de l'hypophyse dans l'oeuf de poule à l'incubateur du 4-e au 20-e jour et au bout de 18 jours il obtenait chez l'embryon une hyperfonction de la glande thyroïde. L'absence de recherches à des étapes moins avancées de développement enlève à ce travail la portée qu'il pourrait avoir.

Afin d'étudier les corrélations endocriniennes aux stades précoces du développement embryonnaire, j'ai entrepris l'examen des réactions spécifiques de l'organisme embryon à divers hormones, à l'aide de la méthode de transplantations chorioallantoides. Dans le présent travail seront communiquées les données concernant l'influence des transplantations du lobe antérieur de l'hypophyse sur la glande thyroïde de l'embryon.

Remarques au sujet des matériaux et de la méthode. On employait pour les transplantations des fragments du lobe antérieur de l'hypophyse de mammifères de la laboratoire adultes: rats, lapins, chiens et poules également. Les fragments préparés pour l'expérience faisaient environ 1 mm³ et étaient greffés sur le chorioallantois d'un embryon de poulet au 7-e ou 8-e jour d'incubation. Au bout de 2 à 10 jours la coquille est ouverte, les greffons et les organes de l'embryon sont fixés pour examen microscopique. Méthode: Helly, Champy, Carnoy, celloi-

dine-paraffine, azine, haematoxyline ferreux d'Heidenhain, haematoxyline alumineux de Delafield.

Développement normal de la thyroïde d'un embryon de poule. L'observation du développement de la glande thyroïde d'après Bradway (2), Sun (9), etc. a montré que les premiers indices de secrétion—l'apparition de gouttes de colloïde—apparaissent à la fin du huitième; les follicules se forment au neuvième jour. Pendant le reste du temps de l'incubation les follicules continuent à se développer jusqu'à la sortie de la coquille. On observe rarement la vacuolisation de la colloïde, l'épithélium folliculaire garde la hauteur des cellules cubiques.

Résultats des expériences: A. Influence de la greffe du lobe antérieur de l'hypophyse des mammifères sur le développement de la glande thyroïde de l'embryon de poule. J'ai décrit les matériaux de cette série d'expériences antérieurement. Les fragments du lobe antérieur, greffés sur le chorioallantois, de même que les tissus de l'embryon, manifestent leur capacité de développement et de différenciation tout en conservant une structure histologique parfaitement normale. La glande thyroïde du récipient répond à l'action de l'hormone provenant du greffon par une certaine accélération de la morphogénèse. Ceci s'exprime avant tout par une intensification du développement des follicules. La hauteur de l'épithélium change peu. La colloïde ne se vacuolise que rarement.

B. Influence de la greffe du lobe antérieur de l'hypophyse de poules adultes sur la glande thyroïde de l'embryon de poule. Au cours des années 1937—1938 j'ai pratiqué plus de 200 greffes de fragments du lobe antérieur de l'hypophyse de poules adultes. Grâce à ces materiaux importants, il a été possible d'étudier la réaction de la glande thyroïde aux diverses étapes de développement et de déterminer le rapport de la réaction avec la quantité de matière transplantée.

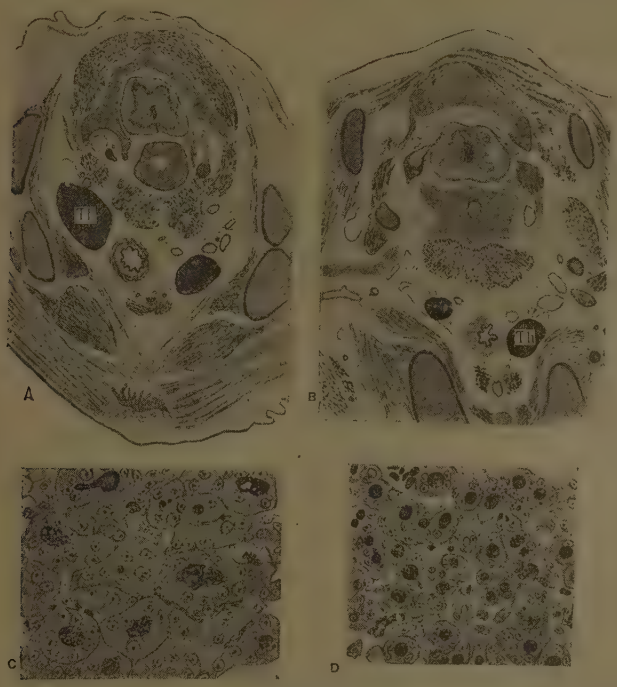
Les fragments du lobe antérieur greffés se conservent merveilleusement. Dès les premiers jours de développement sur le chorioallantois jusqu'aux derniers moments on peut distinguer dans le greffon les formes typiques des cellules du lobe antérieurs, de nombreuses mitoses et l'organisation des cellules en follicules. Dans la plupart des cas le fragment tout entier garde sa structure normale, rarement la partie centrale subit la nécrose.

Pendant les deux premiers jours de développement du greffon, la glande thyroïde réagit par une accélération marquée de la morphogénèse. Chez un embryon de 10 jours soumis à l'action du greffon les dimensions de la thyroïde deviennent sensiblement supérieures à celles d'une glande normale. En même temps, on observe une accélération de la formation des follicules dont les dimensions sont également augmentées.

À la fin du 11-e jour d'incubation, dans certains cas, la glande thyroïde revêt une structure qu'il est difficile de considérer autrement que l'expression morphologique d'une réaction thyréotrope typique. Dans les cas les plus marqués, la structure de la thyroïde excitée est caractérisée par les indices suivants: tout d'abord on est frappé par une hypertrophie brusque de la glande par comparaison à une glande normale de même âge. Tandis que les dimensions d'une glande normale au 11-e jour d'incubation ne dépassent guère 510 × 395 μ sur les coupes, la glande excitée peut atteindre 900×585 μ (fig. 1, *A* et *B*). Le diamètre de certaines follicules d'une glande excitée atteint 30 μ au lieu de 12.8 μ pour une glande normale. Les différences sont encore beaucoup plus tranchantes dans la structure des follicules (figure 1, *C* et *D*). L'épithélium des follicules d'une glande excitée est beaucoup plus haut que l'épithélium folliculaire normal; il atteint 7.2 μ

au lieu de 3.2 μ (moyenne normale). La colloïde au sein des follicules se vacuolise. La glande excitée est fortement hypérémiée et le réseau de vaisseaux qui pénètrent à l'intérieur de la glande revêt le caractère de sinus.

L'étude de la structure de la glande thyroïde des embryons d'épreuve, aux différentes étapes de développement, jusqu'au 17-e jour d'incubation, nous montre que le développement de la glande thyroïde, sous l'influence de



Réaction thyréotrope sous l'influence de la greffe du lobe antérieur de l'hypophyse à un embryon de poule, au 11-e jour d'incubation. *A* et *B*—coupe de l'embryon au niveau de la glande thyroïde (*A*—epreuve, *B*—contrôle). Glande thyroïde (Th)—à l'épreuve—fortement hypertrophiée; *C* et *D*—structure des follicules de la glande thyroïde (*C*—épreuve, *D*—contrôle). A l'épreuve épithélium élevé et colloïde vacuolisée.

l'hormone du lobe antérieur de l'hypophyse, a tendance à subir les modifications suivantes: dans un groupe, le développement de la glande thyroïde se poursuit intensivement, si bien que vers le 15-e ou le 17-e jour d'incubation, lag lande de l'embryon d'épreuve est sensiblement plus volumineuse que la normale, gardant toujours des indices d'excitation. Au 13-e jour, la résorption de la colloïde devient encore plus marquée. On rencontre plus souvent des follicules complètement vides qui se distinguent des jeunes follicules par leur épithélium à hautes cellules. Cependant, cette élévation ne correspond pas toujours à l'augmentation du volume des follicules.

Dans le deuxième groupe, la glande thyroïde montre une certaine tendance de dégradation: les follicules cessent de se développer, leur colloïde ne se forme plus, le nombre des mitoses de l'épithélium folliculaires diminue. On peut supposer que ces phénomènes indiquent un état de dépression de la glande qui succède à un état d'excitation.

C. Modifications que subit l'embryon sous l'in-

fluence de la greffe du lobe antérieur de l'hypophyse. L'influence générale de la greffe sur le développement de l'embryon d'épreuve s'exprime par un retard de croissement. La différence de poids des embryons d'épreuve et de contrôle est peu sensible pendant les premiers jours d'incubation, elle devient très marquée au 14-e 15-e jour. Le retard de développement se manifeste particulièrement chez les embryons dont la glande thyroïde est dans un état de dépression. Dans certains cas, ces embryons présentent des indices de chondrodystrophie, et un développement atypique, du squelette. L'analyse de ces processus fera l'objet d'autres communications. La majorité des embryons d'épreuve présente dans la structure de certains organes internes des modifications qui, parfois, aboutissent à la nécrose. Le plus souvent c'est le foie qui est frappé, des taches de nécrose apparaissent à sa surface; les cellules hépatiques se vacuolisent. Le métabolisme des graisses et des hydrocarbones est troublé. On observe des modifications de structure du mésonéphrose: chez les animaux d'expérience le mésonéphrose est plus volumineux que chez les normaux. Dans certains cas apparaissent des vésicules gonflées de liquide.

Les modifications les plus marquées du système endocrinien s'observent dans la glande thyroïde. Les parathyroïdes subissent également quelques modifications qui seront décrites ailleurs. Dans les autres glandes, un examen superficiel ne revèle aucune modification spécifique.

Conclusions. La question de corrélations endocriniennes, au stade embryonnaire de développement, ne peut être résolue sans avoir etudié la capacité de réactions typiques des glandes endocrines de l'embryon, sous l'influence d'autres glandes endocrines. En effet, si la glande de l'embryon examinée ne réagit pas à l'action spécifique endocrine, c'est que sa différenciation morphologique n'a pas encore atteint le stade voulu de fonctionnement de la glande.

En ce qui concerne la thyroïde de l'embryon de poule, les investigations de différents auteurs ont montré que son activité fonctionnelle commence à la fin de l'incubation[10] ou bien fait défaut en général. Certains auteurs signalent des indices de fonctionnement aux 11-e et 12-e jour d'incubation [2, 9], ce qui n'est pas encore confirmé expérimentalement.

Ainsi, nos données sur la réaction thyréotrope chez l'embryon de poule établissent pour la première fois la capacité de la glande embryonnaire de réagir à l'action de l'hormone du lobe antérieur de l'hypophyse à un stade aussi précoce. On est donc à bon droit de croire que la réaction thyréotrope peut être obtenue non seulement au 10-e jour d'incubation, mais à des étapes moins avancées de développement, lors d'une plus forte action de l'hormone du lobe antérieur de l'hypophyse. Ces données obligent à poser la question du rôle des glandes endocrines sur le développement embryonnaire. Il est probable que pour une série de glandes il faudra reporter le moment, que nous avions l'habitude de considérer comme l'apparition des premiers indices d'activité de fonction, à des stades de développement plus précoces.

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LITTÈRATURE CITÈE

[1] Aron, Rev. française d'endocrinol., 8-e année (1930). [2] Bradmay. Anat. Rec., **42** (1929). [3] Dieterle (1906). [4] Hamett, Am J. Phus.. **82** (1927) [5] Kraus, Endocrinologie, **5**, (1929). [6] Krienecku, C. R. Soc. Biol.. **97** (1927) [7] Landauer, Am, J. Anat., **63** (1929). [8] Loeb a, Basset, Proc. Soc. Exp. Biol. a. Med., **26** (1929). [9] Sun, Phys. Zool., **5** (1932). [10] Ulenhutha. Ebeling. Anat. Rec., **29** (1925) [11] Woodside, Anat. Rec., **67** (7937).

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EXPERIMENTAL MORPHOLOGY

PARATHYREOTROPIC RESPONSE OF THE CHICKEN EMBRYO

By A. N. STOUDITSKY

(Communicated by I. I. Schmalhausen, Member of the Academy, 16.VI.1938)

At the present time there is no doubt as to the stimulating effect of the anterior pituitary lobe on the secretary function of the parathyroid glands. This means that calcium metabolism, which is controlled by the parathyroid glands, stands under an indirect control of the anterior lobe of the hypophysis. There are, however, no exact data regarding these correlations during embryonal development. According to some indications, calcium metabolism at embryonal stages is controlled not by the parathyroid, but by the thyroid gland, since the differentiation of the parathyroid glands begins but at a late period (8). However, it has been established for amphibia that the development of the parathyroid in the tadpole can be stopped by means of hypophysectomy, i. e. that it is correlatively bound with the hypophysis already at an early stage (7). We thus see that the question concerning the mutual relations of the hypophysis and the parathyroid during the development of the embryo cannot be anywise regarded as clear. On the other hand, this question is of great interest, inasmuch as it is involved in one of the most essential aspects of embryonal development—in calcium metabolism. With the purpose of throwing light on the response of the embryonal parathyroid gland to the action of the anterior pituitary lobe, I undertook experiments with the transplantation of bits of the anterior lobe of the hypophysis to the chorion-allantois of the chicken embryo. The results of these experiments are reported in the present article.

Material and methods employed. I studied the influence of the anterior lobe of the hypophysis on the parathyroid glands of the chicken embryo with the aid of transplantation experiments. Bits of the anterior lobe of the hypophysis of adult hens were grafted to the chorion-allantois of 7—8 days old chicken embryos. The effect of such grafts as well as the general response of the embryo to them have been described in a previous report.

The usual methods of histological research were employed for studying the parathyroid glands: Helly, Carnoy, Fleming, Champie, Heidenhain's iron haemotoxylin, Delafield's alum haematoxylin with eosin, asan accerding to Heidenhain, Mallory, saphranin, etc. For the examination of their fat and lipoid components the glands were fixed in Champie's liquid, treated with a mixture of 2% hydrochloric acid and 3% pothassium bichromate for 5—7 days and taken into paraffin. In the preparations, stained with Scharlach-rot and Sudan III, the droplets of lipod stained red and those of neutral fat—black.

Normal development of parathyroid glands in chicken embryo. The histogeny of the parathyroid glands has been studied mainly in mammals. The development of these glands in birds described as early as in 1928 by Verdun[9] has been studied chiefly in its general embryogenetic aspect. According to literary data[1,8], the parathyroid gland of the chicken embryo lacks the morphological symptoms of physiological activity and consists of indifferentiated cords of uniform cells. Histological examination of adult hens[3] also ascribe to the parathyroids of birds a very uniform structure. These data, so different from those available in regard to the histogeny and histological structure of the parathyroid in the mammal, induced me to undertake the study of the normal structure and development of the parathyroid in the chicken.

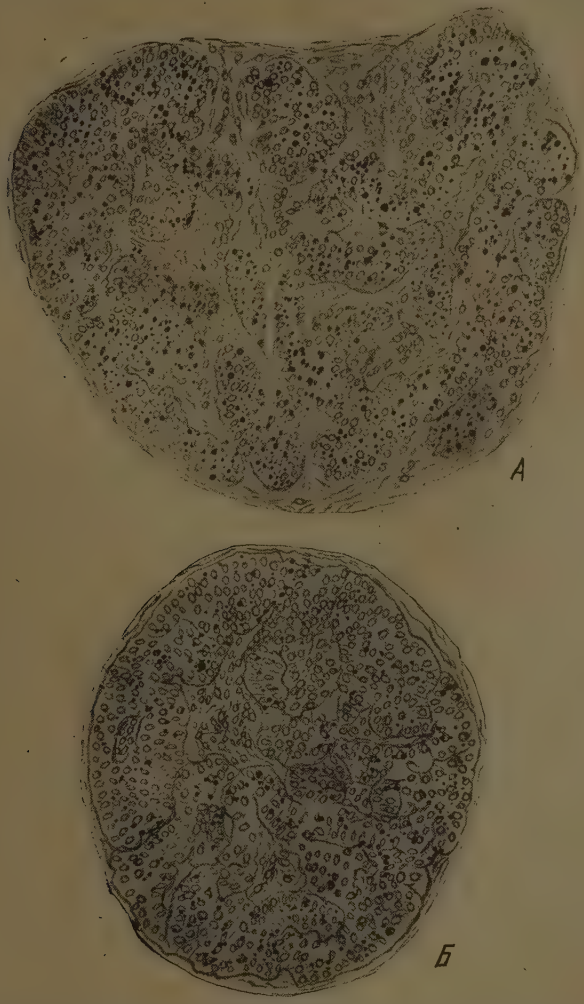
As we know[2], the parathyroid glands of the chicken embryo arise as outgrowth of the 3d and 4th branchial pouches. In the course of the subsequent stages of the embryo's development the embryonic glands begin to descend along the nerve-vascular fascicle of the arteria and vena jugularis, retaining a close contact in their development with that of the thyroid gland and continuing to be bound with the remnants of the branchial pouches by means of long epithelial cords, which run parallel to the nerve-vascular fascicle. As early as by the 8th day of incubation the mutual position of the thymus, the thyroid and the parathyroid glands is the same as the one existing in the adult hen. It follows that the subsequent development of the parathyroid glands is limited to the growth and differentiation of the embryonic glands.

The histological structure of the parathyroid glands of the chicken embryo varies greatly, showing individual variations, and manifests great differences depending on the method used for making preparations. Similar to what is found in the mammal, two main types of cells are encountered here. They may be called, according to Welsch[10], the main ones and the oxiphile ones. The second type includes several variants and their intermediate forms. The extreme variant—the water-clear cell of Norris[4]—is a very large cell with a clear apparently liquefied protoplasm and a vesicle-like nucleus.

Already on the 7th day of incubation cellular cords may be detected in the parathyroid gland, which show the mutual disposition of cells of the same character as in the adult gland. Elongated and pressed against one another, these cells are oriented perpendicularly to the axis of the cord. They keep their polarity which finds its expression in that the apex of the cell with the Golgi apparatus is directed towards the centre of the cord and its base—towards the periphery. On the 8th day oxiphile cells make their appearance among the main cells. The protoplasm of these oxiphile cells is vacuolated and is stained a vivid orange tint by asan. The protoplasm is vacuolated chiefly in the basal part of the cell. On the 10—11th day appear the water-clear cells. A 12-days old parathyroid is already quite typically differentiated.

This early differentiation of the parathyroid glands of the chicken embryo coincides with an accumulation of fat and lipoid inclusions within the cells. The first lipoid droplets are detected in the cells already at the earliest stages of the differentiation of parathyroid glands. By the 7—8th day of incubation in every section, treated so as to make lipoid visible, dozens of vividly stained droplets of varying size may be seen. In a 12—13-days old gland the accumulation of lipoid droplets is still more intense. Lipoid droplets are detected in different parts of the protoplasm of all the three kinds of cells. The accumulation of lipoids continues during the entire period of the embryonal development of parathyroid glands. The adult gland contains not much more lipoid than the embryonal one.

Experimental results. During the first days of the development of the graft on the chorion-allantois the parathyroid gland responds but by an insignificant increase of its rate of development. Comparing 10-day glands after 2—3 days of action of the grafted bils of the anterior pituitary



Parathyreotropic response of the chicken embryo. Accumulation of fat and lipoid inclusions in the protoplasm of the parathyroid gland's cells in a chicken embryo of 15-days incubation under the influence of the grafting of the anterior pituitary lobe (time of growth of the graft—4 days). *A*—test, *B*—control.

lobe, one can see that in some cases the diameter of the test parathyroid is 1 $^1/_2$—2 times the normal size. A particular size increase in manifested by the I corpuscle (parathyreoidea III). Simultaneously a further complication of the gland's structure takes place: the number of cellular cords increases and they turn thinner than those of the control glands. Some 14—16 cords, stretching across the gland, can be counted in a 15-day parathyroid,

3—4 days after the transplantation of the hypophysary fragment as compared with the 6—8 cords of the control parathyroid. An increase of the number of the water-clear and oxiphile cells in relation to that of the main ones proceeds parallel with the complication of the gland's structure.

We should consider as the most interesting effect of the action of the anterior pituitary lobe on the structure of the parathyroid gland, the increased content of fat inclusions in the protoplasm of the gland's cells (Figure). While in a normal 15-day gland fat inclusions are encountered in but a few cells, the parathyroid of the test embryo after 6 days of the graft's action is so enriched in both neutral fat and lipoids that characteristic inclusions can be found in almost every cell of the gland.

After a more prolonged action of the graft (up to 8 days) the parathyroids suffer in some cases alterations which can be interpreted only as phenomena of regression. The size of the glands decreases so much that in some cases the diameter of the test glands is one third of that of the normal glands.

Discussion and conclusions

The intricacy of the structure of the parathyroid gland at its early development stages shows the notion of its being physiologically inactive in the course of its embryonal development([8]) to be erroneous. The complication of the histological structure of the parathyroid gland is setting in by the 7—8-th day of incubation, i. e. simultaneously with the beginning of calcium sedimentation in the embryonal organism. The accumulation of lipoids in the parathyroid gland can be regarded as indicating its increasing physiological activity which agrees with the results of the experiments of Ohmori ([5]) who has proved the functional activity of the parathyroid gland to be connected with the accumulation of lipoids in its cells.

Thus, the morphological pictures observed in the parathyroid gland subjected to the action of transplanted fragments of the anterior pituitary lobe can be, without reservations, interpreted as a rise in its functional activity. The specifity of the described alterations is so obvious that no mistake will be made in designing them as the parathyreotropic response of the embryo.

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25. VI. 1938.

REFERENCES

[1] Landauer, Virchows Archiv, 271 (1929). [2] Lillie, Development of the chicken. [3] Nonidez a. Goodale, Am. J. An., **38** (1927). [4] Norris, Control Embryol., **26** (1937). [5] Ohmori, Okayama Igakkai Lasshi, **41** (1929) [quoted from Ber. wiss. Biol. (1930)]. [6] Rivoir, Presse méd. (1935). [7] Smith, Anat. Rec., **23** (1922); Endocrinology, **7** (1923). [8] Sun, Phys. Zool., **5** (1932). [9] Verdun, C. R. de Biol., **5** (1898). [10] Welsch, Journ. An. a. Phys., **32** (1898).

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EXPERIMENTAL MORPHOLOGY

STUDY OF THE REGENERATION PROCESSES IN THE DEVELOPING BONE

By M. P. AIZUPET

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The regeneration of the bone presents great interest both in its theoretical aspect and for surgical practice. In studying this problem one cannot avoid touching upon such questions as metaplasy, determination and differentiation of cells in the developing bone, the question as to the potency of the individual cell groups etc., the opinion on which is by no means united and regarding which most various, sometimes diametrically opposed views are expressed.

There are many studies and an extensive literature on the regeneration of bones. This literature is mainly surgical and the morphogenetic regularities of this process are chiefly treated therein on human material. A number of researches are available, the authors of which set themselves the task of studying the histogeny of the regenerating bone [Asada([1]), Dietrich([5]), Glaeser([7]), Rhode([9]), Wallis([12]), Wurmbach([15]), etc.). All these authors turned their attention to the histogeny of the bone in its adult condition. As to the regeneration of the cartilaginous and osseous tissues at the early stage of the animals development, this process has not been studied in its details as yet. In particular, the regeneration processes in the embryos of mammals have not been studied at all, although the determination of the potency of cells, and consequently the understanding of the entire process, can be achieved only if we are aware of all the interrelations and potencies of the periosteum in the ontogeny.

The purpose of the present work was to study histogenetically the regeneration of osseous tissue at various stages of the ontogeny in vertebrates.

Tibiae of rat embryos and of new-born rats formed the subject of our study in its first stage. The tibiae of both the rat embryos and the new-born rats were cut across approximately in the middle. The skin covering the dissected bone was sutured. The tibiae operated upon were fixed 1, 2, 3, 4, 5. etc. days after the operation, subjected to histological treatment, and the histogeny of the regenerating bone was studied. The objects were fixed in Zenker's fluid + formol, their histological sections were stained by asan and by the method of Mallory.

The experiments on rat embryos were done on the 18—20th day of gravidity. The operations were made by the intra-uterine method.

At this stage of development an osseous cuff is already formed around the cartilaginous block of the tibia, and the processes of the endochondral

ossification of the cartilage have already begun. On the preparations studied by us we were satisfied that the dissected periosteum of the long bones of the rat embryo regenerated, practically knit, very quickly. A powerful autgrowth of the periosteum, its thickening goes on simultaneously with the knitting. The number of cells in the so-called cambial layer of the periosteum increases at an enormous rate. A cartilaginous callosity is formed between the ends of the dissected bone and around them on the third day after the operation. It should be noted that the bone callosities of rat embryos differ greatly from those of the adult rats, which have been histologically studied and described by Asada([1]), Glaeser([7]) and Dietrich([5]). It should be noted moreover that though the knitting of the dissected tibiae in the embryo is similar to that observed in an adult rat, no such variety of cells is

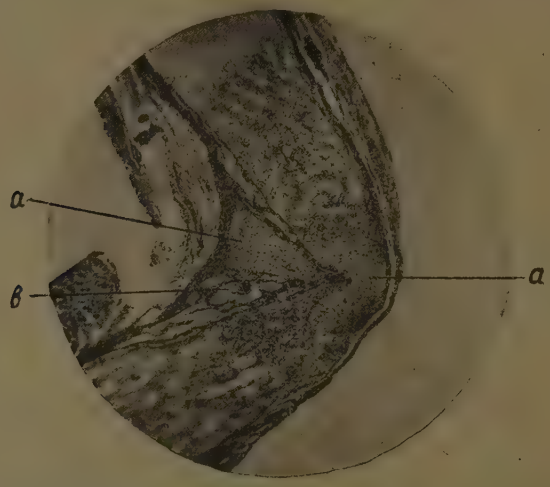

Fig. 1. Bone callosity of the tibia of a rat embryo on the 5th day after the operation: *a*—cartilage, formed at the place of the fracture; *b*—endochondral process in the cartilage callosity, accompanied by a formation of thick bone lamina.

observed in the bone callosity of the embryo, as the one which occurs in the healing bone fractures of adult animals (cartilage, connective tissue, mesenchyme cells). According to our observations, the callosities of embryos are composed of cartilage alone. Remarkably enough, on the third day after the operation processes of endochondral ossification set in inthe already formed cartilaginous callosity. On the fifth day most of the callosity has already acquired the structure of a spongy bone (Fig. 1). A certain interest is presented by the circumstance that the disposition of the osseous lamina in this callosity differs in nothing from their disposition in the bone callosity of an adult animal, i. e. it shows a marked «functional structure». Although at the stage of the rat's ontogeny, at which we experimented on the rat embryo, its extremities do not as yet function as organs of support and locomotion, we see the structure of an organ, adapted for its future function. This indicates that functional stimulation is not the only cause determining the structure of the forming bone.

We have already mentioned that according to our observations a cartilaginous callosity of periostal origin is formed at the spot of fracture both around the ends of the fragments and between them in the healing dissected tibia of the rat embryo. Glaeser([7]) dwells in detail on the origin of the

cartilaginous callosity in the periostal regeneration of the long bones of tritons. He used for his study preparations in which cartilage can be seen lying between the periosteum and the bone centrally of the spot of the operation. The author's opinion is that this cartilage can have been formed only from the periosteum. Fell(6), cultivating in vitro the endosteum of the chicken embryo, observed the transformation of osteoblasts into osteal and cartilaginous cells, She considers that the osteablasts are bipotent. Studitsky(11) is also led to the conclusion that the young osteoblast can differentiate in two directions: it can become either an osteal or a cartilaginous cell.

In regard to the work of the last two authors, the objection may be raised that the cell transformations, taking place in the in vitro cultures, may be absent in the histogeny of tissue in the organism. This was the objection really made.

To solve the question whether the highly differentiated cells of the so-called cambial layer of the periosteum are actually bipotent, we paid serious attention to it in studying our experimental material. In preparations of regenerating tibiae of rat embryos one can see that as early as on the first day after the operation the periosteum in the region of the fracture becomes thicker and the cells of its cambial layer are intensely proliferating. Further one can observe in our preparations the direct transformation of the cells of the cambial layer into cartilage cells. At the initial stage of their transformation these cells have the shape of flat discs. They contain relatively little protoplasm, their nuclei are large; the cells are surrounded by numerous collagenous fibres. Later on these fibres gradually disappear becoming veiled owing to the formation of chondroitin-sulphuric acid, and the cells, experiencing further alterations, are transformed into genuine cartilage cells.

What are the cells of the inner layer of the periosteum? We consider Weidenreich(14), Policard and Leriche(8) etc. to be right, when they say that one can speak of a cambial layer of the periosteum only in regard to the growing bone. The adjacent to the bone layer of the periosteum of an adult animal contains no cambial cells. In adult animals the cells of this layer differ in nothing from the cells of the connective tissue. During the period of growth and development of bones and cartilages the cells of the inner layer of the periosteum receive stimuli, probably of a humoral character, may be partly also of a mechanical one (3, 8, 11). Under the action of these stimulations the cambial layer of the periosteum is formed, from the young cells of which either bone or cartilage is differentiated depending on the stimulus. This formation of both bone and cartilage from periostal cells shows that in the animal organism young cells of a highly differentiated tissue may actually be bipotent.

The process of endochondral ossification of the periostal cartilaginous callosity of rat embryos shows also some peculiarities as compared with what is observed in adult animals. The view commonly acknowledged is that cartilage cells experience regressive alterations in the course of this process. Their basal substance is destroyed, part of the cells, liberated from the cartilage cavities, perish, another part joins the cells of the marrow and becomes indistinguishable from them. The endochondral bone, on the other hand, develops at the cost of osteoblasts of periostal and endostal origin. Some authors, it is true, describe a direct transformation of cartilage cells into osseous ones. Thus, S. Ziba(16) observed alongside of the destruction of cartilage cells and of the formation of osseous tissue at the cost of osteoblasts also the transformation of those cartilage cavities which remain unopened into bone cells. Deineka(4) established that the first layer of the endochondral bone arises to a considerable degree at the expense of cartilage

cells: they are transformed into «the first generation of the osseous cells of the embryonal bone». A. V. Rumjantzew[10] admits that the basal substance arises quite early and experiences a series of ontogenetic transformations of its own, it being possible to indicate three typical consecutive stages of this transformation of the basal substance: connective tissue, cartilage, bone. Our observations are in some respects confirming this view. On the ground of our preparations we establish how the cartilaginous callosity is substituted in the rat embryo and in the new-born rat by the bone tissue, composed of typical bone lamina. Moreover, it can be affirmed with surety that the basal substance of the cartilage takes part in the process. Bands of the basal substance of cartilage, stained a vivid blue, are visible in our preparations (Fig. 2), this being an indication of the unmasking of collagen taking place there. As to the question concerning the cells of the osseous lamina, which develop on the place occupied earlier by cartilage, we leave it open for the present. Two suppositions are possible: either these are cells arisen by transformation of cartilage into bone cells, or they are osteoblasts which have penetrated here from the periosteum. Our next study will be devoted to this question.

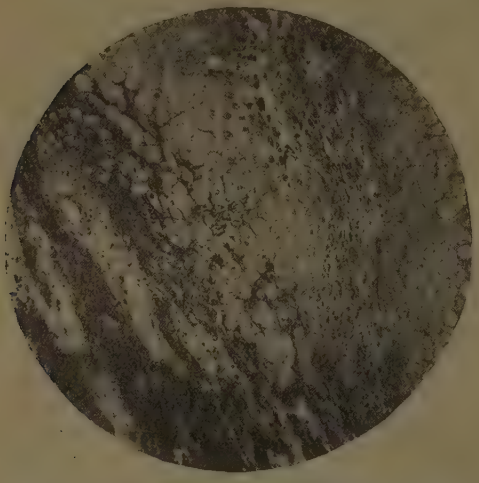

Fig. 2. Formation of bone lamina from cartilage, formed at the place of the fracture of the tibia of a new-born rat.

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Academy of Sciences of the USSR.

Received
25. VI. 1938.

REFERENCES

[1] F. Asada, Arch. f. klin. Chir., **147** (1927). [2] A. Bier, Deutsch. med. Wochenschr., **16** (1918). [3] A. Bier, Arch. f. klin. Chir., **127** (1923). [4] А. Дейнека, Арх. анат., гист. и эмбр., **1**, 1 (1916). [5] H. Dietrich, Arch. f. klin. Chir., **141** (1926). [6] H. Fell, Proc. of Roy. Soc., V (1933). [7] K. Glaeser, Arch. f. mikr. Anat., **75** (1910). [8] Leriche, Policard, Masson et Cilédit (1926). [9] C. Rhode, Arch. f. klin. Chir., **123** (1923). [10] А. Румянцев и А. Березкина, Арх. акад., гист. и эмбр., XVII (1937). [11] Студитский. Известия Акад. Наук СССР (1934). [12] K. Wallis, ZS. Zellforsch. u. mikr. Anat., **6** (1927). [13] K. Wallis, ibid., **7** (1928). [14] F. Weidenreich, Hndb. d. mikr. Anat. d. Mensch., 11/2 (1930). [15] Wurmbach, ZS. wiss. Zool., **129** (1927). [16] S. Ziba, ZS. f. Morphol. u. Anthrop., **13** (1910).

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EXPERIMENTAL MORPHOLOGY

DIFFERENTIATION OF NEUROBLASTS IN CULTURES

By Z. IGNATIEVA

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The question as to the nature of cellular elements in the cultures of nerve tissues remains unsolved up to now. The investigations published during the last decade([1], [2], [3], [4], [6]) display a great variety of opinions concerning the origin of the same cellular forms. Thus, Kapel([2]) assumes that the cells reported to be cultivated may be of nerve origin, but in the course of cultivation they might have gradually differentiated and transformed into epithelial cells. Mihalik([3]) believes all migrating cells in cultures to be macrophagous, or microglial elements. Olivo([1]) and Levi([5]) consider them to be neuroblasts, etc. This difference of opinion is evidently due to the fact that the investigators mostly deal with a great diversity of material, even lacking sometimes precise information as to the part of the nervous system the given material has been taken from.

Considering it possible that various sections of the central nervous system may differ by the character of their growth in culture, we undertook a systematic study of all its principal sections. We started our study from the tissue of the cerebral hemispheres.

The object of the present work was to study the nature of cellular elements and the transformations they undergo in the course of cultivation. Explanting the nerve tissue, taken from embryos of various age, we were trying to establish the potency possessed by the nerve cell at the given developmental stage of the organism. We attach great importance to the solution of these problems, as it would help us to understand many sides of the process of differentiation of nerve tissue which still remain obscure.

Material and procedure. The tissue of cerebral hemispheres was taken from chicken embryos after five days of incubation to hatching, as well as from chickens 1—2 days old. The bits were cultivated by the method of suspension in blood plasm with addition of embryonal extract, diluted with Tyrode fluid in the ratio 1 : 3. No grafting was done. The cultures were left to grow from 20 to 144 hours. As fixatives were used: neutral formole diluted with saline 1 : 4, alcohol 96°, Zenker-formol and the Carnoy-fixative. The preparations were stained with haematoxylin, Toluidin-Blau and Scharlach-Rot. Silvering of the cultures was performed after Bolschovsky.

Original investigations. The study of a vast material convinced us that various cellular elements occurring in the migration zone may be divided into three groups:

1. Small spindle-like cells with a rounded compact nucleus and finely grained protoplasm.

According to some authors([1], [2], [4], [5]) these cells are neuroblasts which being cultivated are capable of differentiation and development of offshoots.

Strange enough, Mihalik does not mention these cells. According to our observations, these cellular forms prove to be especially numerous in the cultures of young embryos after 5—9 days of incubation. Their mass-migration is mostly observed as soon as 24 hours after explantation. It proceeds in the following way: from the bits radiate wide strings contaning many nuclei. These strings run across the zone of liquefaction, which is formed almost directly around the bit, and penetrate the surrounding medium. We failed establishing the boundary of these cells on preparations stained with haematoxylin; after silvering, their spindle-like and rounded shape becomes rather apparent. Stained with silver, the strings assume a yellow-brown colouring, the cells—a brown one.

Having reached the boundary of the liquefaction zone, the cells migrate into the surrounding medium and form rows. Abundantly developing offshoots stretch far off, crossing on their way and forming interlacements. Along with the migration of cells the plasmatic strings thin out and by the end of mass-migration of cells (3rd or 4th day after explantation) they look either like thin stripes with a few cells, or like separate threads. Such plasmatic strings occur sometimes without being accompanied by cells, yet, we have succeeded to retrace their originating from the group of cells lying within the explanted bit. Sometimes, besides, one could see such a thread-like string to run from a cell migrating from the bit, to cross the zone of liquefaction and fuse in the surrounding medium. Upon silvering it is stained yellow. Not unfrequently within a clot of plasm may be met isolated spindle-like cells with several offshoots, among which one is longer than the rest. We have as yet failed to obtain in these offshoots by the silvering method the colouring typical of nerve cells. As to the neurofibrilary apparatus within the cell body, it is easily impregnated and may be seen as thin black threads.

In the cultures of especially young embryos (5—6 days after incubation) such cells frequently display mitosis. Being stained in vivo, they accumulate Neutral-Rot in insignificant amounts. The rod-like chondriosomes are distributed throughout the whole body of the cell.

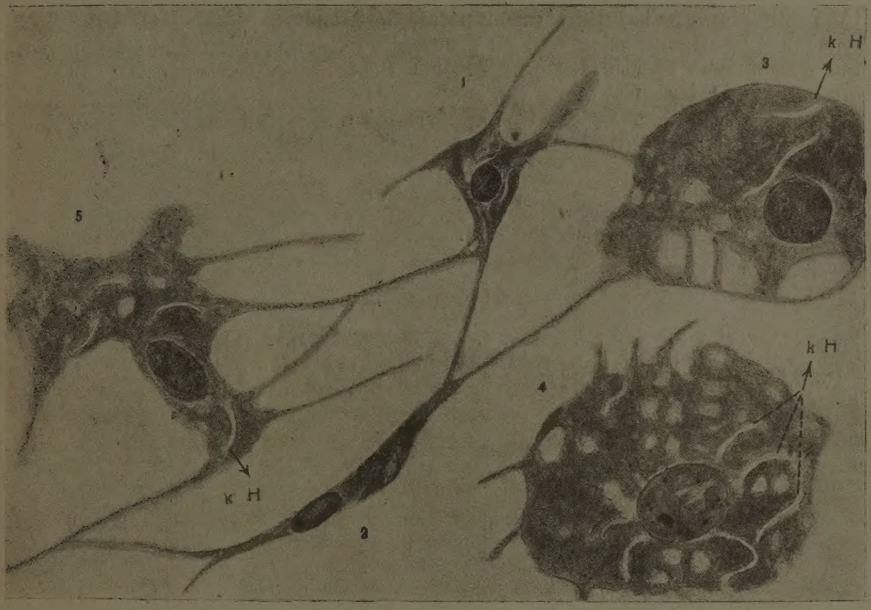
The pathological changes developing in these cells during cultivation closely resemble those frequently observed in nerve cells affected by certain diseases. One of the features characteristic of degenerative processes in the nerve cell we consider the appearance in the protoplasm of Holmgren's canaliculi. Being single at first, they soon cover the whole body of the cell in all directions. This phenomenon is usually accompanied by a considerable vacuolization of the protoplasm. The changes described are reversible only in the case if the nucleus is not involved.

In the cells observed by us all these structural changes may be observed approximately on the 4th day after explantation.

Having taken into consideration all the above-said we arrived at the conclusion that these cells by their nature should be classed among nerve cells. At early stages of the development of the embryo (6—7 days of incubation) they are still undifferentiated and may produce as well macroglial cells as nerve cells. This is shown, first, by the fact that being treated with silver their offshoots do not become typically outlined, and, secondly, by their high degree of viability when cultivated and ability to differentiate, developing offshoots. Though these cells are present in great numbers in the cultures from tissues of older embryos and chickens, no mass-

migration could be observed. They are of a somewhat smaller size, but nuclear structure, shape of the body and developing offshoots are similar to those of the cells described. Most of these cells form a loose layer. In this case we, possibly, have to deal with macroglial elements. When subjected to cultivation, they are more stable than differentiated nerve cells, the structure of which at this time already displays signs of degeneration.

2. Small, isolated round cells. These are found both along the strings and among them. Their body is covered with a kind of radiating thin and short offshoots. They are characterized by rapid fattening. Neutral-Rot is usually accumulated therein in the form of coarse granules. We are yet uncertain as



Culture of cerebral hemispheres of a 17-days old chicken embryo. Growth—96 hours. Fixed with Zenker-Formol. Stained with haematoxylin. Zone of migration. *1, 2*—cells of the first type; *3, 4, 5*—cells of the third type; *kH*—Holmgren's canaliculi in the protoplasm of the nerve cells.

to the nature of these cells, however, the possibility of their belonging to microglial elements seems not to be excluded.

3. Large cells of various shape. Among other migrating elements they abound only in the cultures of older embryos. Their migration begins much later than that of other cells; by that time they already look degenerated. They are characterized by the presence of Holmgren's canaliculi, vacualized protoplasm and the nucleus being shifted to one side. Offshoots are often preserved, though sometimes the cell is quite rounded. When studying sections from cultures of old embryos, one may see that the large nerve cells become rounded as early as in the bit, their protoplasm already exhibiting pathological symptoms. These modified cells migrate from the explanted bit into the surrounding tissue where they assume the most diverse shapes. Put on glass they flatten out and assume the shape of a plate. Most frequently they lie separately, but may occasionally be observed, especially

in live cultures, joint with small spindle-like cells by means of offshoots (figure). Being stained in vivo they accumulate Neutral-Rot in the form of coarse granules, similar to all other degenerated cells.

Further modification of these cells proceed in the following way: the nucleus shrinks and becomes shifted to one side of the body, the vacuoles grow larger and the entire cell assumes the shape of a ring; the offshoots gradually degenerate. We failed to detect neuro fibrilary apparatus in these cells which may, besides, be easily understood, since the neurofibrils are a most sensitive and rapidly degenerating structure. We believe these cells to be differentiated nerve cells, able to survive in cultures but a short time, after which they rapidly degenerate and perish. We thus agree with Kapel as to their being differentiated nerve cells, though we dispose of no evidence that being cultivated they differentiate and become epithelial cells. This is a stage of degeneration and no progressive change. We cannot agree with the conclusions of Mihalik who asserts that all migrating cell forms are macrofagous microglial elements. Although in the course of the degeneration process nerve cells may get the appearance reminding that of macrofages, yet it is possible to discriminate them, bearing in mind the specific characters of an affected nerve cell.

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REFERENCES

[1] O. Olivo, Arch. f. exper. Zellforsch., **4** (1927), **5** (1928). [2] O. Kapel, ibid., **5** (1928); **8** (1929). [3] P. Mihalik, ibid., **17** (1935). [4] P. Lazarenko, ibid., **11** (1931). [5] G. Levi, Erg. d. Anat. u. Entw. Gesch., **31** (1934). [6] S. Mossa, Arch. f. exper. Zellf., **7** (1928).

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